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## MAGNETIC SOUND RECORDING EQUIPMENT

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*The recording of sound on magnetic material is an invention which dates back to the last century. Then it was practically forgotten, left behind by other, more perfected methods of recording; electronic amplifier techniques developed in the meantime have revived interest in it, however. In a relatively short time the magnetic system has now captured a large field of applications which were previously dependent on other methods, besides opening up new spheres in sound recording.*

### Introduction

In 1877 Edison surprised the world with his phonograph, an instrument for the recording and reproducing of sound, in which the vibration set up in a diaphragm by the sound waves were mechanically recorded on a steel cylinder covered with tin-foil. The cylinder was played back by means of the same apparatus. In Edison's phonograph a vertically modulated groove was produced on the cylinder, i.e. a groove the depth of which varied in accordance with the deflection of the diaphragm, in contrast with the gramophone disc introduced by the American Berliner, which is still employed today and in which the groove is modulated laterally, i.e. in the form of a curved line<sup>1)</sup>. There is no need to dwell on the point that the direct, mechanical recording and reproduction of sound was later superseded by indirect methods using an electrical device (an amplifier) between the receiving element (microphone) and the cutter, as also between the needle and the reproducing diaphragm (loudspeaker).

The invention of the sound film introduced an optical method, again in conjunction with the necessary electrical appliances, in which the sound is recorded on the film photographically in the form of a track, either of constant width and variable density, or constant density and variable width; the reproduction is obtained by means of a scanning light slit and photo-electric cell.

In between the two systems of the gramophone record on the one hand and the sound film on the other, there is the Philips-Miller system which employs a sound track with variable-width modulation, cut mechanically in a celluloid tape coated with black lacquer and scanned by optical means<sup>2)</sup>.

Each of these methods has its own specific advantages and drawbacks, and each is employed today in its own particular sphere. In recent years, however, another system of sound recording and reproduction, based on the use of magnetic materials, has found wide application. As long ago as 1898, Valdemar Poulsen, a Dane, patented a magnetic system of recording, the sound being recorded on a steel wire wound on a drum and passed between the poles of an electromagnet. Naturally his equipment was primitive seeing that, at that time, the amplifier had not yet been invented. Poulsen's system was not developed and was accordingly shelved. About in 1935, however, when modern techniques of amplification were known, Poulsen's idea was taken up by Stille, Schüller and others in Germany; Lorenz in that country and Marconi in England placed a machine that worked with a steel tape on the market. The speed of the tape was about 1.5 m/sec, and it was possible to record frequencies up to 6000 c/s; the tape was wound on a spool about 70 cm in diameter and the playing time was a good half-hour.

<sup>1)</sup> See L. Alons, Philips techn. Rev. 13, 134, 1951 (No. 5).

<sup>2)</sup> See Philips techn. Rev. 1, 107-114, 135-141, 211-214, 231-236, 1936.



As early as 1927 F. Pfleumer in Germany and Joseph O'Neill in the U.S.A. had developed paper and plastic tapes coated with a layer of magnetic material. In Germany, in the period 1940-1945, magnetic recording systems using such tapes were further developed on a large scale.

Magnetic recorders marketed nowadays are either of the "wire" or "tape" variety. The wire recorder will not be discussed here. Certainly, the equipment is usually less complicated and therefore cheaper than the tape recorder, but it has certain definite disadvantages; if the wire breaks, or if it is inexpertly handled, it will often become hopelessly tangled; the joining of two ends, either to repair a break or for the "montage" of different selections, can only be achieved by knotting, which is a much less reliable method than the splicing of two ends of tape. Several other, more fundamental, disadvantages of wire in contrast with tape are mentioned in the following paragraphs.

#### General construction of a tape recorder; properties of the tape

The tape is magnetized by means of a "recording head" which is in effect an annular electromagnet with small air-gap, along which the tape is passed (fig. 1). The modulation current flows in the magnet coil; this current may consist of the alternating current generated by a microphone as a result of the sound to be recorded. For the reproduction the alternating magnetization produced in the tape, the "sound-track", is scanned by a "playback" head, which is almost identical with the recording head; it contains a coil, across the ends of which an alternating voltage is produced by the passage of the tape, this voltage being amplified for passing to a headphone or loudspeaker.

The material on which the sound is to be recorded must be capable of permanent magnetization and

must therefore exhibit a certain hysteresis and remanence; for reasons which will be explained presently, moreover, the coercive force must not be too low, say in the order of  $\mu_0 H_c = 0.01 \text{ Wb/m}^2$ <sup>3)</sup>. Lastly, the structure of the material must be very finely crystalline. This requirement is related to the inherent tape noise, which must always be as little as possible. Magnetization of the tape does not vary continuously from point to point, but is constant in magnitude and direction within certain small zones (Weiss zones) and changes suddenly between the boundaries of one zone and another. The noise arising from this discontinuity is so much the less according as the Weiss zones are smaller, and, since one such zone can at most extend to the confines of one whole crystal, a finely crystalline structure is obviously desirable. The simplest course is to use as basis a powdered magnetic material. As a rule, iron oxide, either red ( $\gamma\text{-Fe}_2\text{O}_3$ ) or black ( $\text{Fe}_3\text{O}_4$ ) in a grain size of less than  $1 \mu$  is employed.

The powder can be incorporated in the tape in two different ways: it can be either mixed thoroughly with a thermoplastic material such as polyvinyl chloride, the tape being then extruded or rolled in the hot condition; or it can be mixed with a lacquer which is subsequently applied as a very thin layer (e.g.  $15 \mu$ ) to a paper or plastic tape about  $35 \mu$  in thickness.

The low noise level resulting from the use of finely crystalline material is also the reason why tape is usually preferred to wire. Certainly a solid wire can be made that will exhibit very small crystals, but wire introduces another very important cause of noise which is entirely absent in tape and which will be explained in the following.

A homogeneous tape, made in accordance with the first of the two above-mentioned methods, is the most able to meet the numerous and, in some cases stringent, mechanical requirements imposed by its use in magnetic recording instruments. The chief requirements, standardized in some countries, are that the tape shall not break under loads up to 2 kg, and that the elongation one minute after removal of a load of 1 kg shall not exceed 0.2%. Further, the tape must not pull out of straight, assume a curved cross-section or shrink noticeably; it must be easily cemented and sufficiently smooth to enable it to slide over the recording head and other components without causing undue wear. The tape itself should not be subject to too much wear either; in particular it must shed as little dust as possible, must be flexible and maintain all these

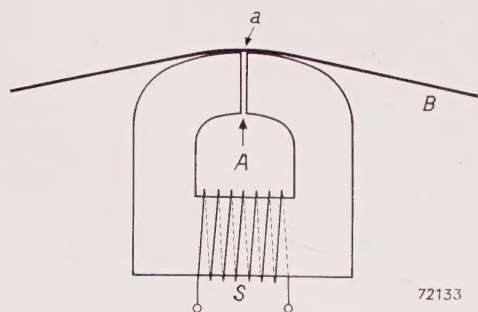


Fig. 1. The recording head, over which the magnetic tape passes. The modulation current flows in coil S. The tape B passes in a gentle curve over the gap A. Recording takes place at a.

<sup>3)</sup> In C.G.S. units;  $H_c = 100$  oersted.



characteristics against the effects of humidity and tropical temperatures.

As far as wear is concerned, it will be clear that even the least, usually unavoidable, wearing of the tape would soon fill up the air-gap in both recording and playback heads with magnetic powder and thus provide a magnetic shunt for the tape or coil. This difficulty is surmounted by filling the gap in advance with a non-magnetic material. Such material must be harder than that of which the pole pieces of the heads are made, to prevent it from wearing away and leaving an open gap. Beryllium copper is very suitable for this purpose.

Standardized dimensions for the magnetic tape are: thickness maximum  $60\ \mu$ , width 6.20-6.35 mm. The length is unspecified, but is usually not more than 1000 metres. The tape is magnetized either across the full width, or only for a part of it, to enable two different tracks to be recorded next to each other, or for stereophonic recordings, in which case two tracks differing only very slightly from each other are made simultaneously. In special cases, as many as six tracks have been made on one tape.

#### Recording with high-frequency auxiliary field; the response curve

If a magnetic particle in the tape passes the alternating magnetic field of the recording head (i.e. the gap and its immediate vicinity) so rapidly that the transit time is short compared with the periodicity of the field (i.e. of the sound to be recorded), it may be assumed that this magnetic field is constant during the time that the particle is passing through it. The particle is magnetized to a certain degree, mainly in the longitudinal direction of the tape<sup>4</sup>) and a part of this magnetization is retained in consequence of the remanence; the remanent magnetization at any point in the tape is of course more pronounced according as the magnetic field is stronger at the moment when that point is passed.

This is not enough, however. It is necessary for the magnetization of the tape to be directly proportional to the field strength, since otherwise the well-known disturbance due to non-linear distortion (overtones and combination tones) will be manifested in the sound reproduced. Although the magnetization curve of most magnetic materials, including that of magnetic tape, is anything but linear, the required linearity can be obtained by an artifice.

This consists in superimposing on the modulation current an alternating current ("auxiliary current"), the frequency of which is well beyond the audible range. The way in which the auxiliary high-frequency field thus produced accomplishes its purpose will not be discussed here; this will be explained in an article shortly to be published in this Review, dealing with the mechanism of magnetic recording in greater detail. It is sufficient to say here that, in principle, the necessary linearity can also be ensured by means of D.C. pre-magnetization; this introduces a considerable amount of background noise, however, which is absent when the high-frequency method is used.

The frequency of the auxiliary field must be so high that it changes polarity several times during the passage of a magnetic particle through it. On the other hand, it has just been pointed out that the field at the sound frequency may be regarded as constant during that time. Naturally this is no longer the case either where the highest audio frequencies are concerned. What we then record at a given point on the tape represents, as it were, an average of the field strengths to which the particle is subjected, not when passing the gap, as might be supposed, but when traversing a small zone at the edge of the gap. For a further explanation, reference must once more be made to the article announced above. One result of this formation of an average is that the high frequency components of the sound are recorded relatively weakly.

A similar effect is encountered in the playback. The alternating voltage induced in the coil of the playback head by the movement of the tape is proportional to the speed of variation of the flux in the coil. This flux is dependent on some kind of "average" of the magnetization over a length of the tape equal to the width of the air-gap. An average is thus formed in this case, too, again with detriment to the higher frequencies in the sound. The smaller the air-gap and the greater the tape speed ( $v$ ), the higher the frequency at which this effect will become noticeable; the air-gap must be smaller than the shortest wavelength to be recorded and reproduced by the tape. The wavelength of a sinusoidal signal of frequency  $f$  is  $v/f$ . At the much used speed of  $v = 76.2\text{ cm/sec}$  (30 inch/sec) and with  $f = 10\,000\text{ c/s}$ , the wavelength  $v/f = 76.2\ \mu$ . For the speed mentioned the playback head is usually given a gap of  $10\ \mu$  to  $20\ \mu$ ; the gap in the recording head may be made slightly larger, say  $25\ \mu$  to  $40\ \mu$ , this being advantageous from the point of view of more uniform magnetization throughout the thickness of the tape.

<sup>4</sup>) In principle it is also possible to magnetize the tape in the direction of the width or thickness. With the usual longitudinal magnetization a certain amount of magnetization always occurs in the direction of the thickness as well.



Let us now consider the response curve of the magnetic recorder as a whole, viz. the alternating output voltage as a function of the frequency of a tape recorded with constant current amplitude. Since the voltage induced in the playback coil is proportional to the speed of variation of the magnetization, it increases with frequency; in other words the response curve exhibits a natural gradient to a factor of 2 (6 dB) per octave. In the higher frequencies the increase is not so marked, and at the upper end of the range there is even a drop. In part, this is due to the above-mentioned influence of the finite dimension of the gap, but there are other causes, too. In the sound track the shortest wavelengths tend to be levelled out, seeing that closely spaced elements, magnetized with opposed polarity, have a demagnetizing effect on each other; in principle this effect should be counteracted by the iron in the playback head as soon as the relevant section of the tape approaches it, but, owing to the small intervening space that always remains between tape and head, this counterbalancing effect is not complete. (In order to suppress the effect in point as much as possible, the coercivity of the magnetic material should be high.) In the shortest of the wavelengths concerned this space between tape and head provides an opportunity for the lines of force to close up to a large extent without passing through the magnetic circuit of the playback coil. Lastly, the effect of eddy current losses in the recording and playback heads becomes more pronounced as the frequency is increased.

Whereas the response curve of the magnetic recorder accordingly assumes the form shown in *fig. 2*, a flat characteristic is our ultimate aim.

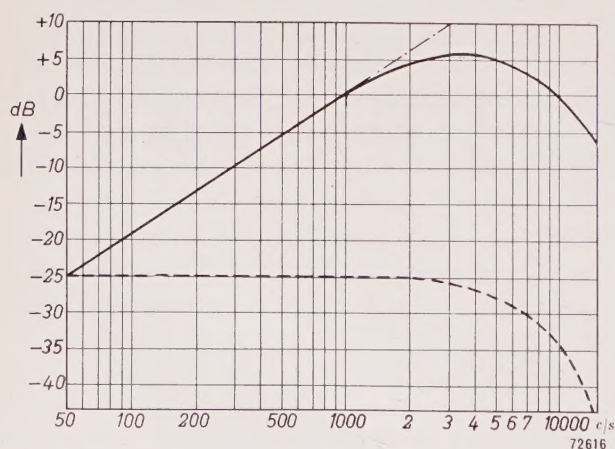


Fig. 2. Overall response curve of a magnetic recorder, showing on a relative scale the alternating voltage obtained by reproduction with constant amplitude in the recording coil, as a function of the frequency. The full line shows the response without any correction; the broken, for the greater part horizontal, line is obtained with an RC network in the playback amplifier. The higher frequencies still require boosting.

Our first course in pursuing this aim is to correct the rise in the middle frequencies. This can be achieved by means of a simple RC network in the playback amplifier, the result being the largely horizontal broken line in the figure. In the second place, the higher frequencies need to be boosted. This can be effected in both the recording and the playback amplifier. In principle it is better to boost the high notes at the recording stage, as they are then placed further beyond the noise level of the tape. This does involve the risk, however, that, in the louder passages the saturation magnetization of the tape will be reached in the higher frequencies, with consequent distortion. A compromise is therefore made, the correction being introduced partly in the recording and partly in the playback. It is a point in our favour that in orchestral music — during the recording of which the interference due to noise and distortion is the most disturbing — the strength of those components of which the frequency is high is usually less than in the middle and lower frequencies. This circumstance makes it possible to shift the greater part of the necessary correction in the higher frequencies towards the recording side.

### Erasing

Recordings on magnetic tape exist by reason of the magnetized condition of the particles in the tape, and they can, therefore, be modified by a strong magnetic field, in contrast with other types of recording which, apart from the effects of physical damage, are permanent. For this reason magnetic tape recording should not be stored in places where strong magnetic fields set up by permanent magnets, transformers, motors etc. occur. The requirement that magnetic tapes should not be too sensitive to magnetic fields with which they are accidentally brought into contact is yet another reason why the coercivity of the material should not be too low (see above).

At the same time, however, the susceptibility of magnetic recording to magnetic influences is put to good use in that used tapes can be rendered magnetically neutral and thus made serviceable for fresh recordings. This erasing is effected by a separate "erasing" head, this being in the main identical with the recording and playback heads. A high-frequency magnetic field is produced in the gap of the erasing head, such that the material of which the tape is made, when passing the head, is subjected to a number of hysteresis loops which first increase and then decrease to zero on the *B-H* curve. After passing the erasing head, then, the tape is once more in the demagnetized condition.



In most recorders the sequence of operations is such that, for recording purposes, the tape first passes through the alternating field of the erasing head, thus ensuring a perfectly "clean" basis for the recording.

In the case of wire recorders it is much more difficult to erase a recording completely. This is due to the fact that the wire is drawn through slots in the recording and erasing heads, and that side of the circular cross-section of the wire which faces the open end of the slot is only slightly exposed to the recording and erasing fields. Since wire is always more or less subject to twisting, after erasing, certain parts of the wire will, generally speaking, retain traces of the previous recording. After repeated recordings and erasures all these residues tend to produce a background babble, which constitutes extra noise, not occurring in tapes.

heads and induction from external magnetic fields at mains frequency (hum). The gaps in the recording and playback heads must be precisely at right angles to the tape; any obliqueness would again affect the higher frequencies<sup>5)</sup>.

### The tape driving mechanism

For a satisfactory quality of reproduction the speed of the tape passing the recording and playback heads must be highly constant. The mechanical problems to be solved in order to meet this and certain other requirements connected with the tape drive are worthy of being discussed<sup>6)</sup>. In this we shall base our considerations mainly on the design of Philips' tape recorders.

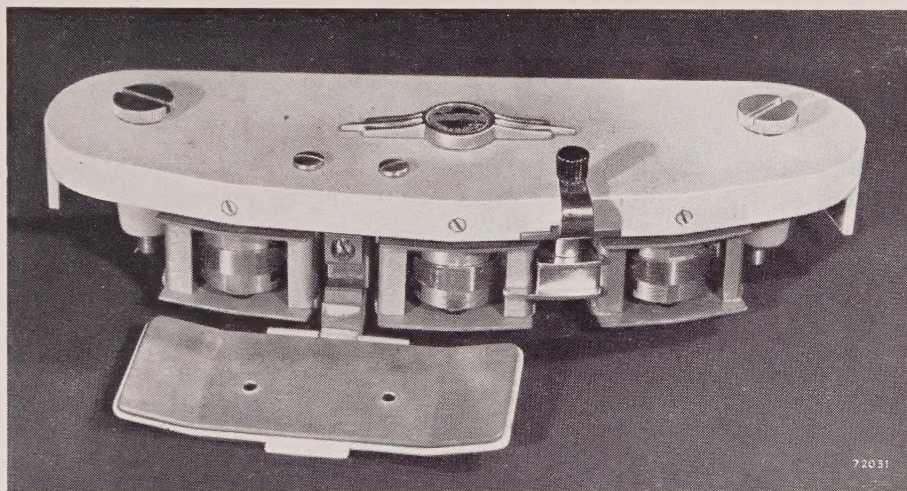


Fig. 3. The magnetic heads of a Philips tape recorder, arranged as a single unit (see fig. 6); from right to left the erasing, recording and playback heads. In front of the two last-mentioned heads a mu-metal screen is provided, this being placed over the tape to eliminate interference due to cross-talk and possible alternating fields in the vicinity.

The frequency of the erasing field is normally 35 or 40 kc/s and the auxiliary recording field generally somewhat higher, e.g. 100 kc/s. The larger magnetic recorders will usually incorporate two separate oscillators to provide these frequencies, but other equipment may have only one oscillator generating the two frequencies required, as harmonics of a fundamental frequency.

In the larger equipment the erasing, recording and playback heads are mounted in a common, detachable, holder. Fig. 3 shows the unit as fitted to one of Philips' tape recorders, this assembly being fitted with plugs for direct connection to the amplifiers. The playback and recording heads are housed in compartments of high permeability material such as mu-metal, in order to limit cross-talk between the

In principle, uniform motion of the tape is obtained by means of an accurately machined spindle which is driven by a synchronous motor and against which the tape is pressed by a rubber roller (fig. 4).

<sup>5)</sup> Actually, it is only necessary for the gaps in the recording and playback heads to be parallel to each other. In order to ensure proper reproduction of a recorded tape on any suitable instrument, perpendicular gaps should of course be used. — It may be added here that, in the case of "amplitude" recordings, as in the Philips-Miller system as well as in one of the photographic methods, oblique positioning of the gaps results in non-linear distortion of the sound: see J. F. Schouten, Philips techn. Rev. 6, 110-119, 1941. In our own case (intensity recording) there is only linear distortion (variation in the response curve, as mentioned in the text).

<sup>6)</sup> See also J. J. C. Hardenberg, The transport of sound film in apparatus for recording and reproduction, Philips techn. Rev. 5, 74-81, 1940, in which article similar problems concerning the drive of the Philips-Miller system are discussed.



This driving spindle must of course be made of non-magnetic material and should be as little as possible subject to wear.

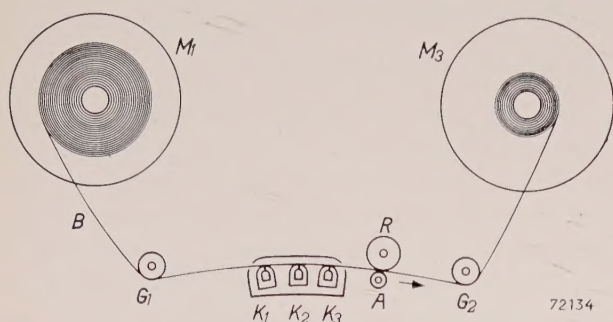


Fig. 4. Principle of the tape drive. *A* = transport spindle; *B* = magnetic tape; *R* = rubber roller holding the tape against the transport spindle. Erasing, recording and playback heads *K*<sub>1</sub>, *K*<sub>2</sub> and *K*<sub>3</sub> are assembled in a common unit. *M*<sub>3</sub> = winding spool; *M*<sub>1</sub> = supply spool. *G*<sub>1</sub> and *G*<sub>2</sub> = guide rollers.

Eccentricity is reduced to a point below the tolerance limit of about 4  $\mu$  by grinding the spindle in its own ball-bearings. The rubber roller must be highly uniform in composition and the pressure such that the tape will just be prevented from slipping when a certain tensile force (about 800 gr) is applied.

Synchronous motors ensure the roller to be rotated at a perfectly constant speed as an average, but there is always a certain slight oscillation superimposed on this rotation, owing partly to the finite number of poles in the motor and partly to a natural tendency to "hunt"; in accordance with the load on the motor, the rotor lags in phase with respect to the stator field, and this phase angle, generally speaking, does not adjust itself aperiodically, but in an oscillatory manner. This tendency is suppressed as much as possible in the design of the motor, a discussion of which would take us too far afield, however. Any residual effects, including irregularities in running due to the poles, are further reduced by placing a flywheel on the transport spindle and by transferring the motion from the motor by means of a flexible coupling ("mechanical filter").

When recording or reproduction is to be commenced, the transport spindle must quickly assume its correct working speed, viz. within a matter of seconds. In order to ensure this, and to permit the motor to come easily into synchronism with the mains, the motor must have a certain starting torque; this will have to be so much the greater according as the inertia of the flywheel is increased. It is therefore not advisable to use a heavier flywheel than is necessary to keep fluctuations in speed just within the tolerance, for which limits have been standardized internationally; the motor can then be suit-

ably proportioned to ensure that the correct speed is obtained within the desired space of time (in Philips magnetic recorders about 2 seconds).

The Comité Consultatif International de Radiodiffusion have laid down the following limits for fluctuations in the speed of magnetic recorders intended for broadcasting purposes: max. 1 per thousand at 30 inch/sec tape speed, max. 1.5 per thousand at 15 inch/sec, max. 5 per thousand at 7.5 inch/sec. Special equipment has been developed by means of which, when reproducing a recorded frequency of 3000 c/s, a direct reading as a percentage can be obtained of variations in speeds, if desired split into fluctuations of frequency < 20 c/s (wow) and > 20 c/s (flutter).

The synchronous motor driving the transport spindle must not fall out of synchronism when the mains voltage fluctuates within the extremes of the prevailing values. In some equipment, moreover, it must be suitable for two tape speeds, in which case the dimensions of the flywheel should be such that speed fluctuations will not be too great at the lower speed, the motor being such that, with this (relatively heavy) flywheel, it will develop the proper speed quickly enough for the higher of the two tape speeds as well<sup>7)</sup>.

During recording and playback, not only the transport spindle has to be driven, but also the spool on which the tape is wound up. Simultaneously the supply spool from which the tape is being unwound must be braked in order to ensure that the tape remains taut; this is essential in order to keep the space between the tape and the heads, as mentioned above, sufficiently small (fig. 4). For rewinding purposes the supply spool has to be driven, and, to prevent looping of the tape, the other spool is then braked. As a rule, higher speeds are required for rewinding than for recording of playing; it is also convenient to be able to run the tape through faster in the forward direction when collecting particular passages of recordings for montage purposes.

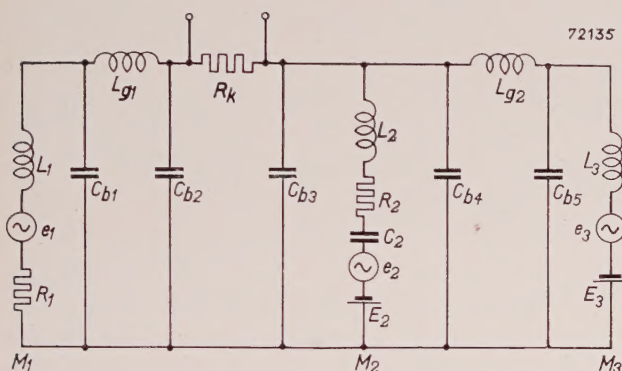
In the larger magnetic recorders all these conditions are met by providing separate motors for the two spools. For recording and playing both are energized with a fraction of the mains voltage; the working spool then slowly winds up the tape and the emptying spool receives a slight torque in the oppo-

<sup>7)</sup> In recorders that work with a slow tape speed the motor can be run with the tape disengaged from the transport roller; the tape can then be started up immediately at any given moment by springing it suddenly against the roller. If the tape speed is on the fast side, however, this involves too much risk of snapping the tape; for that matter, even in slow-speed instruments suitable means have to be provided for absorbing such jerks.



site direction, so that the tape is held taut. For rapid running or rewinding the motor is fully energized to give a tape speed some 10 or 20 times faster than normal.

The tensile forces exerted on the tape by the motors and the inevitable slight fluctuations in these forces must obviously affect the uniformity of the tape movement past the recording and playback heads. The outcome is a complex system of forces, the effect of which can best be studied with the aid of an equivalent electrical circuit. This is shown in *fig. 5*, a brief explanation being given in the caption.



*Fig. 5.* Equivalent electrical circuit of the mechanical tape drive, by means of which the influence of the various components on the speed of the tape can be investigated. A rotating mass is represented by a choke ( $L$ ), a constant force by an e.m.f. ( $E$ ), fluctuations in the force by an alternating voltage ( $e$ ), a friction or a constant braking force by a resistor ( $R$ ) and a spring force by a capacitor ( $C$ ). In particular,  $M_2$  = synchronous motor (represented by  $E_2$  and  $e_2$ );  $L_2$  = flywheel with rotor and transport spindle;  $C_2$  and  $R_2$  = spring force and internal friction of the flexible coupling;  $M_3$  = motor for winding spool (with  $E_3$  and  $e_3$ );  $L_3$  mass of rotor and winding spool;  $M_1$  motor of supply spool (with braking forces  $R_1$  and  $e_1$ );  $L_1$  = mass of rotor and supply spool;  $R_k$  = friction of tape against the heads;  $L_{g1}$  and  $L_{g2}$  mass of guide rollers;  $C_{b1}$ ,  $C_{b2}$ ,  $C_{b3}$ ,  $C_{b4}$  and  $C_{b5}$  = resilient sections of the tape.

The tape speed at the head is represented by current flowing in  $R_k$ ; in the ideal case this would be a direct current without any ripple. To this end the ripple voltage from  $e_2$  must be fully smoothed by  $L_2$ ,  $C_2$ ,  $R_2$  and  $C_{b3}$ ;  $e_1$  is largely smoothed by  $L_1$ , the remainder is smoothed by the "filter"  $C_{b1}$ ,  $L_{g1}$ ,  $C_{b2}$ ; similar conditions apply to  $e_3$ .

The forces applied by the spooling motors (regarded as constant for the purposes of the diagram and represented by the e.m.f.  $E_3$  and resistance  $R_1$ ) will in general depend on the diameter of the roll of tape wound, or left, on the respective spools. In principle, this could result in a variation in the tape speed from the beginning to the end of the tape. It is possible to design the motors so that the forces will be independent of the diameter; experience has shown, however, that a certain amount of variation in the force, to the extent of say 1:4, is quite permissible. For such a limitation of the variation it suffices to make the spool cores large enough.

It should be noted here that the type of drive just described, in which constant speed is ensured mainly by the transport spindle, can be used only in tape recorders, not with wire. In wire-recording, variations in speed are therefore usually appreciably greater, and a gradual change in the speed from the beginning to the end of the recording is by no means so easily prevented. This is particularly troublesome when different parts of a recording are extracted and assembled; during the playback a sudden change in the pitch of the sound may then be audible.

Rapid spooling and rewinding of the tape introduces another problem of its own, namely the possibility of stopping the tape suddenly when a wanted passage is found. Both the spools must then be capable of stopping quickly and simultaneously if the tape is not to be allowed to run slack, or to be stretched possibly to the breaking point. Numerous devices have been tried out. In Philips' recorders this problem is solved by means of carefully balanced brake-bands on the spools, actuated by springs which are released by Bowden cables when a push-button is depressed to stop the motors. This system is very reliable, besides being superior to, and more flexible than the electromagnetic braking system often employed, which necessitates special measures to avoid audible interference due to induction effect in the tape. To prevent unnecessary wear, a slide is pushed up when the tape is to be run through quickly, thus holding the latter clear of the erasing and recording heads (this can be seen between the heads in *fig. 3*); the tape, however, maintains light contact with the playback head, so that the sound can be more or less followed and a wanted passage rapidly located.

### Some current models of tape recorder

Of the different models of magnetic recorder made by Philips let us now consider two with a view to explaining some of the special features, viz. a large model for professional use in broadcasting studios etc. (type No. 10028/03) and a small portable "semi-professional" set, which is at the same time suitable for amateur use (type No. EL 3540).

The first of these is depicted in *fig. 6*. There are two tape speeds, viz. 76.2 and 38.1 cm/sec which, with a standard tape length of 970 m corresponds to a playing time of 21 and 42 minutes respectively. The response curve is flat to within plus or minus 2 dB between 30 and 15 000 c/s and the noise level (background, hum etc.) is 54 dB below signal level on full output, i.e. the level at which distortion is 2%. Variations in the tape speed, both transient and throughout the length of the tape, are within 0.1% at the speed of 76.2 cm/sec and within 0.15% at 38.1 cm/sec.



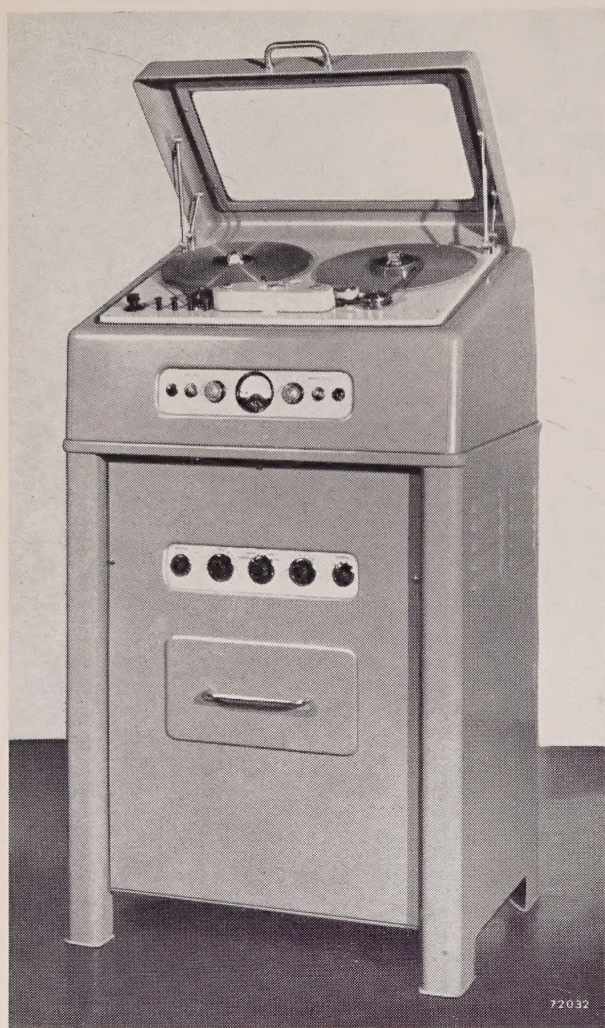


Fig. 6. Large magnetic recorder for broadcasting and film studios, etc., type No. 10028 03. On the top panel will be seen the tape spools, the head assembly and the control knobs. The oscillators and voltage supply units are housed inside the cabinet.

The entire equipment is contained in a steel cabinet. The driving mechanism is mounted under the top panel, on the upper side of which will be seen the two spools, the recording, playback and erasing unit and the press-button controls (see also *fig. 7*). The cabinet is fitted with two drawers which slide out to the rear and contain the recording amplifier, the playback amplifier with correcting network, the two high-frequency oscillators (100 kc/s for recording and 40 kc/s for erasing) and a common voltage supply unit. The drawers are equipped with built-in plug pins to facilitate their removal and replacement for servicing purposes.

To ensure that the whole unit shall be as versatile as possible, the high-frequency and speech currents are both variable; by means of two switches (lower panel, *fig. 6*), each of these can be adjusted to any one of 11 stages. A control is also provided for

boosting the high tones to a variable extent during recording, so that optimum playback quality can be obtained whatever the characteristics of the magnetic tape used.

The upper front panel contains a meter from which can be read the values of the recording and playback currents, as well as the high frequency and erasing currents. The recording and playback amplifiers are designed to receive and deliver the signal at "line" level, i.e. a level that is suitable for transmission by telephone line to a broadcasting station. A jack in parallel with the meter serves for connection of headphones for monitoring the recording or playback.

Owing to the very fine tolerances employed in the workmanship, the sensitivity of the recording and playback heads — as well as their other characteristics — vary very little. Defective heads can, therefore, be replaced without any readjustment, to give the same output level within about 1 dB.

*Fig. 8* depicts the other model under review. This is of much simpler design and has only two heads, one for erasing and one for recording and playback; a single oscillator (pre-magnetization and erasing are both effected at 45 kc/s); only one motor, which



Fig. 7. Close-up of the top panel of the magnetic recorder depicted in *fig. 6*. The components of the driving mechanism sketched in *fig. 4* are easily recognized. At the front of this panel, on the left, will be seen the push-buttons for the control. The first of these (left to right) is for the rewind; rotation of the knob on the extreme left controls the speed, both forward and reverse. The second button starts the recorder at normal speed forward; this is the "playback" setting, seeing that the playback head is permanently connected to the amplifier (all amplifiers are of course previously switched on). The third push-button brings in the erasing head and recording head. To avoid the possibility of erasing part of a recording by depressing the button accidentally, this button is interlocked in such a way that it functions only when depressed together with the second button (which merely starts the machine). The fourth button returns all the others to neutral and stops the instrument by braking the motors.



by mechanical means fulfils all three of the functions of normal drive, fast forwards and fast rewind; finally, recording and reproduction are both carried out by means of the same amplifier: one output pentode is used for both functions by means of a switching manipulation, however, on the understanding that for the playback the pentode of the oscillator, which would otherwise be inoperative, is connected in push-pull with the first-mentioned valve in order to further improve output quality.

Amplifier distortion at 1 W output is below 2%; at 6 W output below 5%. This set is equipped with an electronic indicator for checking modulation depth of modulation, and the noise level in the reproduced sound can, if desired, be further reduced at the expense of the higher tones, by means of a tone control. With a view to the use of this instrument by amateurs, its operation (loading and unloading etc.) has been kept as simple as possible.

A special feature which we should mention in



Fig. 8. Small portable magnetic tape recorder type No. EL 3540 for professional or amateur use. At the left the microphone; at the right the loudspeaker. The recorder itself is housed in one small carrying case.

By thus making the most of the limited number of components, an instrument is obtained that ensures quite good reproduction and, besides, offers all the usual operational features. As the amplifier includes an output stage, the instrument is suitable for use with much lower input voltages and at higher output voltages than in the large model; hence, if desired, the speech current can be recorded direct from a microphone and the tape can be played back direct through a loudspeaker. This tape recorder works with a speed of 19 cm/sec and the spools hold about 515 m of tape, giving an uninterrupted playing time of 45 minutes. The response curve is flat to within plus or minus 3 dB between 60 and 6000 c/s and the noise level is 50 dB below signal level on full drive. Transient variations in tape speed do not exceed 0.15% and, although the over-all variation is rather more than in the larger unit, it is still on the low side, viz. about 0.5%.

closing this article is the device for demagnetizing the recording/playback head. Each time the set is switched over from recording to playing, a current impulse is produced across the coil in the head, which results in a certain, albeit weak, remanent magnetization. The same thing happens if too strong a signal overloads the head during recording, or if a steel object, say a screw-driver, is brought too close to the head. As remanent magnetization of the head introduces distortion and noise in the recording, it is essential to demagnetize the head. In the semi-professional tape recorder under review this is effected automatically each time the change-over switch is operated; when the button is depressed the recording/playback head is temporarily disconnected and placed in contact with a charged capacitor. This capacitor discharges a damped oscillation through the head, thus fully demagnetizing it, regardless of the extent of the spurious magnetization.



In professional tape recorders with separate recording and playback heads, an impulse of this kind is not produced, so that there is less need to guard against other sources of magnetization, occasional demagnetization in the usual way by means of a decaying alternating field being sufficient.

**Summary.** A description of the principle of modern magnetic-tape sound recording. The composition of the tape is discussed from the point of view of the magnetic and mechanical

requirements to which the tape has to conform; this is followed by a general review of the manner in which the recording is made, using an auxiliary high frequency field, the overall response curve and its correction and the erasing of the recorded sound track. Details of the tape drive are given, in particular the means of ensuring very constant speed for both recording and playing, as also for rapid running and rewinding for "montage", checking and so on. In conclusion two of the tape recorders made by Philips are reviewed, viz. a large unit for use in broadcasting studios etc., and a small portable instrument suitable for both professional purposes and amateur use. The more important features of these two apparatus are mentioned, together with some special details of the design.

## A STABILISED EXTRA-HIGH TENSION RECTIFIER FOR 5000 V, 50 mA

by P. PERILHOU † \*) and J. CAYZAC \*).

621.314.67:621.316.722.1

*A direct voltage that at the maximum variations in load varies only to the extent of roughly 0.01%, that is subject to fluctuations equal to not more than 1/3000th part of the mains voltage variations and that is variable within wide limits; these are the outstanding features of a stabilised rectifier designed specially for laboratory work in the field of decimetric and centimetric waves.*

There are many instances in which a rectifier is found preferable to a battery (with charging equipment), since it demands no regular maintenance. Moreover, especially when high voltages are required, a rectifier is usually very much less voluminous and cumbersome and is therefore more easily transportable. Nevertheless, if a particularly constant voltage is needed, rectifiers have this disadvantage that the direct voltage is subject to relative variations equal to those inherent in the mains supplies to which they are connected, unless special precautions are taken to prevent this. Added to this is the fact that the internal resistance of the conventional type of rectifier is usually higher than that of a battery of comparable output. The output voltage of a rectifier is thus generally more dependent on the load than is the case with batteries.

Stabilised rectifiers which are capable of meeting this difficulty were introduced considerable time ago; these include a control circuit that will maintain a sufficiently constant output voltage on a fluctuating mains supply and/or load to an even

greater extent than a battery working on varying loads. More than 10 years ago this Review included an article on the subject of circuits suitable for this purpose <sup>1)</sup>, the principle of which is illustrated with reference to fig. 1.

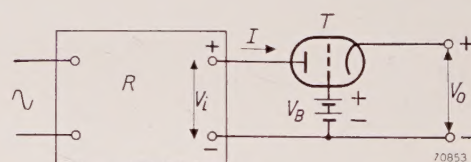


Fig. 1. Circuit for producing a stabilised direct voltage  $V_0$ .  $R$  = rectifier with smoothing filter, delivering non-stabilised voltage  $V_i$ .  $T$  = control valve.  $I$  = direct current.  $V_B$  ( $\approx V_0$ ) = constant reference voltage supplied by a battery.

In this figure  $R$  represents a rectifier with smoothing circuit. The output voltage  $V_i$  is dependent on the mains voltage and the load current  $I$ . The control circuit comprises a triode  $T$ , known as the control valve, which functions as a variable resistor in series with the load, in the positive D.C. line. A battery of which the voltage  $V_B$  is roughly equal to the desired output voltage  $V_0$ , is connected between the grid of the valve and the negative line, the polarity being such that the difference  $V_B - V_0$  serves as grid bias.

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<sup>1)</sup> H. J. Lindenhovius and H. Rinia, A direct current supply apparatus with stabilised voltage, Philips techn. Rev. 6, 54-61, 1941.



Briefly, the working is as follows. The voltage  $V_0$  finds a level that is slightly higher than  $V_B$ , such that the grid is just sufficiently negative with respect to the cathode to allow the required current  $I$  to flow. Increases in  $V_0$  raise the negative potential of the grid and, conversely, decreases in  $V_0$  reduce it, so that any variation in  $V_0$ , whatever the cause of it may be, is at once counteracted. A simple calculation will show that the attenuation factor  $\alpha$ , that is, the ratio of the relative variation in  $V_i$  to that in  $V_0$ , is nearly enough equal to the amplification factor  $\mu$  of the control valve (see the article quoted in note <sup>1</sup>). This amplification factor may be, say, 20.

Very similar considerations apply to the internal resistance, i.e. the quantity that indicates the extent to which  $V_0$  varies in consequence of changes in the strength of the current. As shown in the article referred to in note <sup>1</sup>, the internal resistance at the output terminals is roughly  $\mu$  times lower than that of a rectifier without control circuit.

The higher the amplification factor  $\mu$ , the less the dependence of the output voltage on the input voltage and on the current. The amplification factor of triodes is, however, limited to values which are not particularly high. In order to make the stabilisation as accurate as possible, it is therefore customary to employ a direct voltage amplifier ( $A$ , fig. 2) to amplify the difference between the battery voltage  $V_B$  and the output voltage  $V_0$  by a factor  $n$ . The output voltage is then used as grid bias for the control valve; instead of  $\mu$ , we now have  $n\mu$ , the attenuation factor is  $n$  times higher and the internal resistance  $n$  times lower than without the amplifier. It is a simple matter to obtain a value of  $n$  of 100, which represents a considerable improvement compared with the original conditions.

The position is less satisfactory, however, if the output voltage  $V_0$  is so high that it would be considered a disadvantage to have to use a sufficiently large number of batteries to ensure that the reference voltage  $V_B$  would actually approximate to  $V_0$ . It is then preferable to work with a lower value of  $V_B$  and to compare this lower value with a fraction ( $p$ ) of  $V_0$  (fig. 3). This means sacrificing some of the accuracy of the compensation, viz. to the extent of  $1/p$  and, in order to restore the original degree of accuracy, the amplification must be increased by  $1/p$ .

If the voltage  $pV_0$  is taken from a variable potentiometer in the manner shown in fig. 3, the output voltage  $V_0$  becomes continuously variable, for  $V_0$  will in every instance adjust itself so that  $pV_0$  is approximately equal to  $V_B$ .  $V_0$  is

thus varied by adjusting  $p$ . (In principle the same purpose is served by varying  $V_B$ ; for continuous control, however, it would then be necessary to connect a potentiometer across at least a part

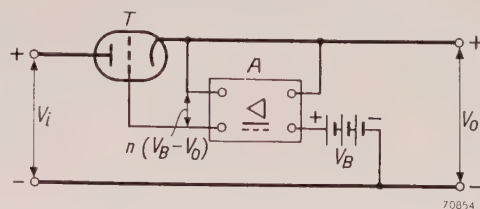


Fig. 2. Circuit as in fig. 1, but with direct voltage amplifier  $A$  to amplify the small difference in potential  $V_B - V_0$  by a factor  $n$ .

of the battery, and this part would be under constant discharge; hence this is not a very suitable method, especially when dry batteries are used. It may be noticed that in figs 1, 2 and 3 the battery supplies no current.)

So much, then, for the principle on which the voltage stabilisation is based. Further details and variations, as well as a description of a stabilised rectifier for 300 V, 100 mA, will be found in the article referred to in note <sup>1</sup>.

Special circumstances may necessitate special solutions. For example, if an unusually high direct voltage and only a small amount of current are required — which means a very high load resistance — stray capacitances have a detrimental effect. A solution for a problem of this kind has already been described in an earlier issue of this Review<sup>2</sup>) (acceleration voltage, for an electron microscope, of 50 to 100 kV, at only a fraction of a milliampere). In the following a particular instance will be described, viz. that of a direct voltage

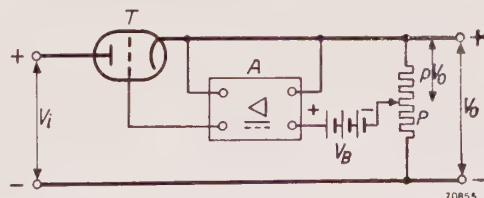


Fig. 3. As fig. 2, but with  $V_B$  approximately equal to the voltage  $pV_0$  tapped from  $V_0$  by means of a potentiometer.

source required for laboratory work in the field of decimetric and centimetric waves. The velocity modulation valves employed to generate these frequencies are such that the frequency is dependent on the voltage supplied to the valve. As the frequency must be very constant indeed, variations

<sup>2</sup>) A. C. van Dorsten, Philips techn. Rev. **10**, 135-140, 1948.



in the supply voltage must be kept to an extreme minimum. The special feature of this problem lies mainly in the amplification circuit, which, owing to the high direct voltage, differs somewhat from the conventional arrangement.

#### Main requirements to be fulfilled by the rectifier

The following are the conditions imposed on a stabilised direct voltage source of the kind envisaged:

- 1) A direct voltage continuously variable from 900 to 5000 V.
- 2) Direct current: maximum 50 mA.
- 3) Relative fluctuations in the direct voltage to be at most 1/1000th part of the relative mains voltage fluctuations, which may amount to 10% of the nominal value.

triode TB 2.5/300<sup>3</sup>); the rectifier proper consists of a transformer  $Tr_1$ , two rectifier valves and a smoothing circuit. Reference to the amplifier will be made later.

As an output voltage of 5000 V is required, and as about 1000 V is lost in the control valve, the voltage  $V_i$  at the output of the smoothing filter should be at least 6000 V. Apart from the required direct current of maximum 50 mA the rectifier should be capable of supplying an extra 15 mA for feeding the amplifier, the potentiometer  $P$  (and other resistors, the purposes of which will be mentioned presently), as well as a voltmeter shunted across the output terminals.

The required flexibility between 900 and 5000 V is obtained by means of a control in large and small

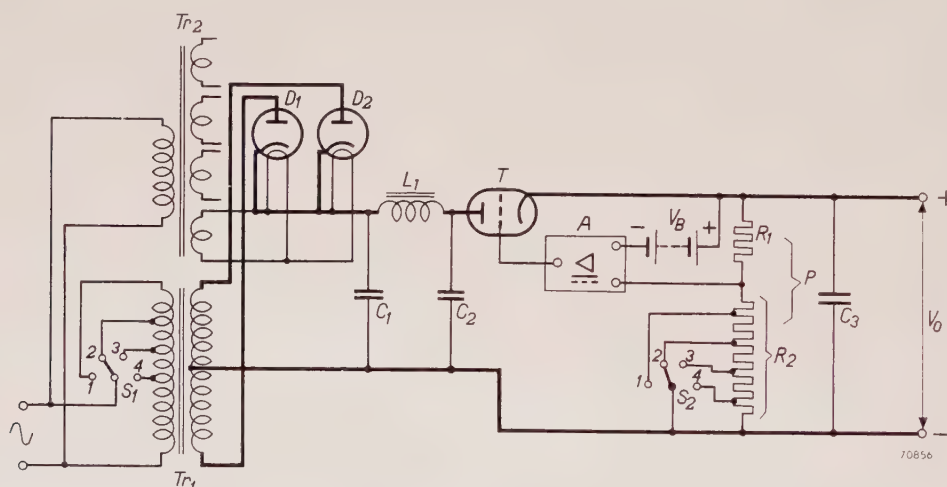


Fig. 4. Circuit of stabilised rectifier for 5000 V, 50 mA,  $Tr_1$  = H.T. transformer,  $S_1$  tapping switch for adjustment of the secondary voltage of  $Tr_1$  in four stages.  $Tr_2$  = heater transformer.  $D_1$ ,  $D_2$  = diodes.  $C_1$ - $L_1$ - $C_2$  = smoothing filter.  $P$  = potentiometer comprising fixed resistor  $R_1$  and tapped resistor  $R_2$  with switch  $S_2$  (ganged to  $S_1$ ).  $C_3$  = output capacitance.  $V_B$ ,  $A$  and  $T$  as in the previous circuit.

- 4) At 5000 V, variations of 50 mA in the current must not result in voltage variations greater than 0.5 V (0.01%); in other words, the internal resistance (for D.C.) should be less than 10  $\Omega$ .
- 5) The internal impedance to frequencies of some hundreds of kilocycles must not exceed 20  $\Omega$  (at 5000 V).
- 6) On 5000 V the combined noise and ripple voltages must not exceed 10 mV.

The methods by which these conditions are met are described below.

#### Description of a stabilised rectifier to deliver 5000 V, 50 mA

Fig. 4 shows the main details of the circuit. The control circuit is similar to that depicted in fig. 3, the control valve being the transmitting

steps, combined with continuous adjustment. Switch  $S_2$  (fig. 4) provides the rough adjustment by varying the resistance of  $R_2$  in stages.  $R_2$  and a fixed resistor  $R_1$  together constitute the potentiometer  $P$ ; when  $S_2$  is operated the fraction  $p$  tapped from the voltage  $V_0$  and, with it, the voltage  $V_0$  itself is varied, seeing that  $pV_0$  is at all times approximately equal to the constant reference voltage  $V_B$  (approx. 300 V).

In order that the control valve will not have to absorb a potential difference of several kilovolts when a low value of  $V_0$  is required, say 1000 V, the secondary voltage from the transformer  $Tr_1$  is modified simultaneously with the rotation of switch  $S_2$ , by means of switch  $S_1$  which is ganged

<sup>3</sup>) Described in Philips techn. Rev. **10**, 273-281, 1948, in particular pp. 278-280.



to it. For the four positions of these switches,  $V_0$  assumes values between the following limits:

$S_1$  and  $S_2$  in position 1:  $V_0$  between 900 and 2100 V,  
 $S_1$  and  $S_2$  in position 2:  $V_0$  between 1900 and 3100 V,  
 $S_1$  and  $S_2$  in position 3:  $V_0$  between 2900 and 4100 V,  
 $S_1$  and  $S_2$  in position 4:  $V_0$  between 3900 and 5100 V.

As will be shown presently, the arrangement is rather more complicated, seeing that when the switches are operated, not one but three resistances have to be varied.

Control in small stages (200 V) as well as the continuous control (over a range of at least 200 V) are effected by smaller tappings on the resistor  $R_2$ , and by varying a portion of  $R_2$ , respectively. These two methods of control are not shown in fig. 4.

### The amplifier

The question now arises: what must be the amplification  $n$  of the amplifier  $A$  to ensure that relative mains voltage fluctuations appear in the direct voltage  $V_0$  1000 times reduced?

We have seen that  $n\mu$  is equal to the desired attenuation factor  $a$  if the battery voltage is taken to be equal to  $V_0$ , but that, with a lower battery voltage, the gain must be  $1/p = V_0/V_B$  times greater. To this must be added a factor  $V_i/V_0$ , to make allowance for the fact that a certain change in  $V_i$  will represent  $V_i/V_0$  times as great a percentage of  $V_0$  as of  $V_i$ ; hence:

$$n\mu = a \frac{1}{p} \frac{V_i}{V_0} = a \frac{V_i}{V_B}.$$

With  $a = 1000$ ,  $V_i = 5000 + 1000 = 6000$  V and

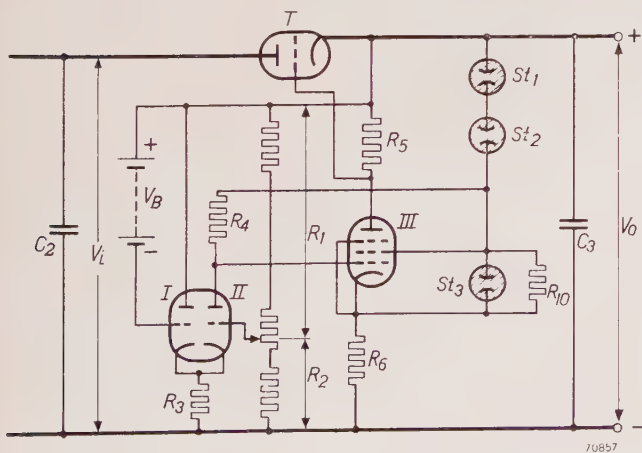


Fig. 5. Circuit diagram showing some of the details of the direct voltage amplifier  $A$  in fig. 4. Section I of the double triode ECC 40 works as a cathode follower and section II as first amplifier valve. The second valve, III, is a pentode EF 42.  $R_3$  and  $R_6$  are biasing resistors.  $R_4$  and  $R_5$  are anode load resistors.  $St_1$ ,  $St_2$ ,  $St_3$  = stabiliser tubes type 85A1, supplying anode voltage for triode II and screen voltage for pentode III.  $R_{10}$  is referred to later (fig. 13). Other references as in fig. 4.

$V_B = 300$  V, it follows that:

$$n\mu = 20\,000.$$

As the amplification factor  $\mu$  of the triode TB 2.5/300 is 20,  $n$  must be 1000.

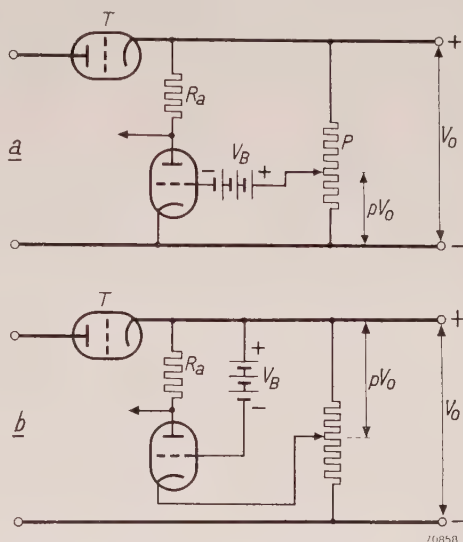


Fig. 6. a) Circuit of the first valve of the amplifier  $A$  (fig. 4). This circuit is practicable only when  $V_0$  does not exceed a couple of hundred volts. b) In this case the first amplifier valve is on the positive side; the lower end of the potentiometer  $P$  now serves as a biasing resistor and produces negative feedback, which may be strong enough to reduce the gain in this stage to about unity.

This cannot be achieved with a single stage of amplification; two stages are necessary and these yield a gain of 3000; hence the condition that  $a$  must be at least 1000 has been amply fulfilled.

The first stage (fig. 5) comprises a triode (section II of a double triode ECC 40), and the second a pentode (EF 42), the amplification factors of these valves being 25 and 120 respectively. The other half of the ECC 40 serves as an input stage with no gain but ensuring freedom from certain difficulties.

If  $V_B$  is to be constant, the battery must be connected in such a way that it will supply no current, i.e. in series with the grid of a valve. This valve could be the first in the amplifier — connected in the manner shown in fig. 6a — if the voltage  $V_0$  were not more than two or three hundred volts. If  $V_0$  is much higher, however, it becomes difficult to provide a direct voltage coupling between the first valve (the electrodes of which are roughly at the potential of the negative line) and the second amplifier valve, which has to deliver the grid bias for the control valve and is therefore roughly at the same potential as the positive line.

This can be overcome by adopting the method illustrated in fig. 6b, in which the first valve is also placed on the positive side. The negative end of the potentiometer  $P$  now functions as a biasing resistor and also introduces a certain amount of negative feedback, which reduces the gain. When  $V_0$  is high, this resistance is so high that the gain of the first



valve is reduced approximately to 1. Two further stages are thus necessary for the required gain of 1000.

In the suggested circuit (fig. 5) there is an input stage without gain, followed by two stages of amplification, with this difference as compared with fig. 6b that the two sections of the ECC 40 constitute a Miller compensator. The feature of this circuit is that variations in the cathode emission (e.g. due to fluctuations in the heater current caused by mains voltage fluctuations) are largely compensated<sup>4)</sup>. For this reason special stabilisation of the heater current is unnecessary.

Section I of the ECC 40 works as a cathode follower: the potential of the cathode follows that of the grid, i.e. the potential  $V_0$ , by way of the battery. When  $V_0$  increases, for example, the potential of the cathodes of this valve rises by the same amount, whereas the grid potential of section II increases only slightly. The anode of II and the control grid of the EF 42 thus acquire a higher potential, the current flowing in the load resistor  $R_5$  rises, the grid of the control valve becomes more negative and thus counteracts the increase in  $V_0$ . In this way, then, variations in  $V_0$  have the effect of adjusting the bias of the control valve in the right direction.

The extreme values which this bias must be able to assume are  $-7$  and  $-170$  V and the anode current of the EF 42 must be capable of variation between corresponding limits.

The amplifier is fed from the direct voltage terminals, and three stabiliser tubes type 85A1<sup>5)</sup> in series ensure that the voltages supplied to the amplifier valves are constant.

To simplify assembly, the circuit is so arranged that no high potential differences with respect to the positive line occur at any point in the amplifier itself; the cathodes of the ECC 40 are at about  $-300$  V and the cathode of the EF 42 carries about  $-255$  V with respect to positive. The remainder of  $V_0$  is taken up by the resistors  $R_3$  and  $R_6$  (fig. 5).

#### *Adjustment of the direct voltage*

It has already been stated that the direct voltage can be controlled in large and small steps as well as continuously, the main control in the range from 900 to 5100 V being in four overlapping steps of 1200 V. This is achieved by means of switches  $S_1$  and  $S_2$  (fig. 4), which provide four values of the secondary voltage from the transformer  $Tr_1$ , and four potentiometer ratios  $p$ . Simultaneously with  $p$  (i.e. resistor  $R_2$ ), the resistors  $R_3$  and  $R_6$  must be suitably varied, for which purpose switches  $S_1$  and

$S_2$  are ganged to  $S_3$  and  $S_4$  (fig. 7),  $R_3$  and  $R_6$  being so varied that the correct amount of current flows in each of these resistors (4.45 mA, which also applies to  $R_2$ ). A fifth switch,  $S_5$ , also ganged to  $S_1$  and  $S_2$ , sets the voltmeter across the output terminals for a measuring range of 4000 V in positions 1 and 2, and for a range of 6000 V in positions 3 and 4.

A second set of ganged switches,  $S_6$ ,  $S_7$  and  $S_8$ , provides adjustment of the voltage in six steps of 200 V.

Further,  $R_3$ ,  $R_2$  and  $R_6$  can be varied by means of ganged, continuously variable resistors, giving a gradual control over a range of 200 V. In this way any required direct voltage between 900 and 5100 V is at once obtainable.

#### *Prevention of instability*

As will be seen from fig. 2 onwards, the amplifier and control valve form a closed circuit. Under certain conditions this circuit is liable to oscillate. For example the following undesirable situation might arise when the H.T. transformer is switched on. Before the switch is closed, no current flows through  $R_5$  (fig. 5), the bias of the control valve is therefore zero and the resistance of this valve (anode to cathode) is low. Consequently, when the circuit is made, the direct voltage  $V_0$  rises rapidly, particularly if the load is inductive. The current flowing through  $R_5$  then also rises, although, when there is some delay in the amplifier, less quickly. It may then happen that, when  $V_0$  reaches the required value, the grid bias (across  $R_5$ ) will not have acquired its corresponding value, with the result that  $V_0$  rises beyond the appropriate level, to the point where the bias becomes so highly negative that the control valve forms a high resistance.  $V_0$  then drops, but because of the delay in the amplifier the potential across  $R_5$  arrives too late to be effective and  $V_0$  is reduced too much, and so on. Instead of attaining equilibrium,  $V_0$  thus fluctuates about the required value.

It would take us too far afield to enter into all the details of the precautions necessary to prevent such instability, and only brief mention of these will be made.

Everything possible has been done to reduce delaying influences, i.e. phase displacements in the amplifier; hence the absence of coupling capacitors and grid leaks. In this way the frequency at which instability might set in is shifted towards the higher values (beyond 15 000 c/s). A choke ( $L_2$ , fig. 7) having a high impedance at that frequency, is connected in series with the control valve.

<sup>4)</sup> S. E. Miller, Sensitive D.C. amplifier with A.C. operation, Electronics 14, 27-31 and 105-109, Nov. 1941.

<sup>5)</sup> T. Jurriaanse, A voltage stabilizing tube for very constant voltage, Philips techn. Rev. 8, 272-277, 1946.



(In order to prevent this choke from increasing the internal impedance too much, it is shunted by a 10 000  $\Omega$  resistor). A capacitor  $C_3$  across the output terminals serves to check abrupt variations in  $V_0$  (and also ensures a low internal impedance at high frequencies, viz. always less than 20  $\Omega$ ).

The potentiometer  $P$  consists of wire-wound resistors; these are not non-inductive, and the

at the output to the grid and cathode of triode *II* without phase displacement, and with a polarity such that the fluctuations are largely compensated. It is therefore not objectionable that a fairly high ripple is present on the output side of the smoothing filter or, in other words, this filter ( $C_1$ - $L_1$ - $C_2$ , figs 4 and 7) need not conform to any special requirements. With the values of  $C_1 = C_2 = 0.5 \mu\text{F}$  and  $L_1 = 40 \text{ H}$

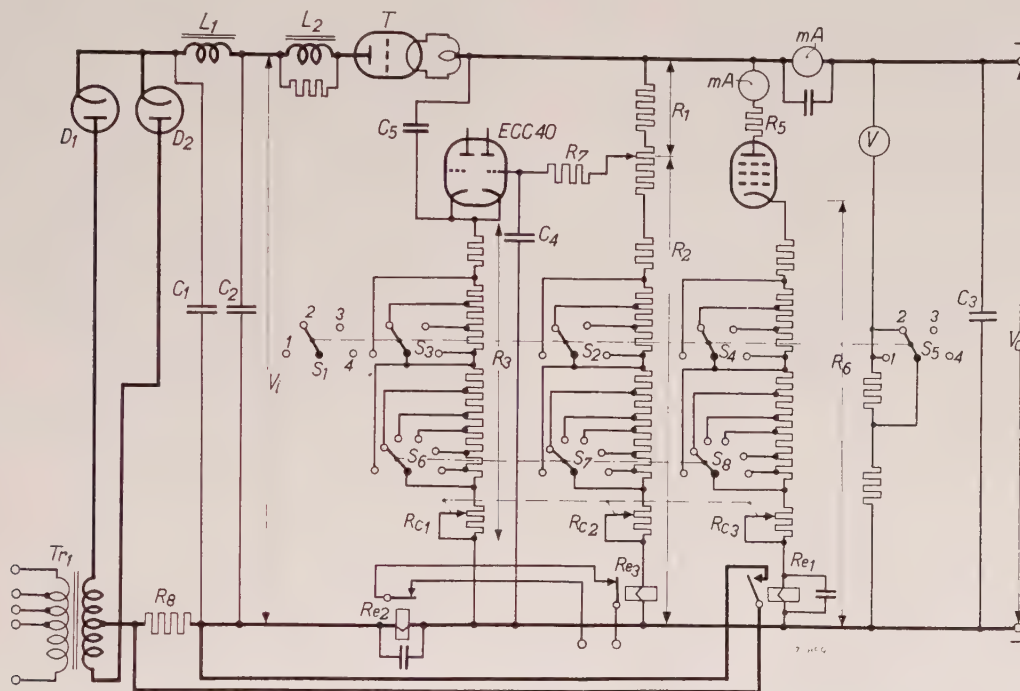


Fig. 7. Ganged switches  $S_1, S_2$  (cf. fig. 4),  $S_3, S_4$  provide adjustment of the voltage  $V_0$  between 900 and 5100 V in stages of 1200 V. Each such stage is divided into further steps of 200 V by means of ganged switches  $S_6-S_7-S_8$ . Within each of these smaller steps  $V_0$  can be varied continuously by the ganged variable resistors  $R_{C1}, R_{C2}, R_{C3}$ . Switch  $S_5$  (ganged to  $S_1-S_2-S_3-S_4$ ) sets the voltmeter  $V$  to a test range of 4000 V in positions 1 and 2, and 6000 V in positions 3 and 4.

To prevent instability:  $L_2$  (8 H), output capacitor  $C_3$  (0.2  $\mu$ F), decoupling circuit  $R_7$ - $C_4$ .

**Protective measures.** To avoid transient current surges in the rectifier valves when the generator is switched on: resistor  $R_8$  (subsequently shorted by relay  $Re_1$ ). Overload protection: relay  $Re_2$ . Excess voltage protection: relay  $Re_3$ . When  $Re_2$  or  $Re_3$  opens, the H.T. transformer circuit is broken (see fig. 8).

Other references as in previous diagrams.

potentiometer thus tends to produce a phase displacement which favours instability. The decoupling circuit  $R_7$ - $C_4$  (fig. 7), however, prevents any alternating voltage component in the voltage tapped from  $P$  from reaching the grid of triode  $II$ .

These precautions are sufficient to suppress any tendency towards instability.

### The smoothing

The control circuit ensures that not only random, but also regular fluctuations in the input voltage — the ripple — appear in the output in an attenuated form. The capacitors  $C_4$  and  $C_5$  (fig. 7) pass possible alternating voltage components

as employed by us, the ripple across  $C_2$ , with  $V_0 = 5000$  V and at full load, is about 30 V. The corresponding ripple (including noise) at the output is 1/3000 of this value (10 mV).

### Protective devices

Rectifier valves with oxide-coated cathodes suffer damage if anode current is allowed to flow before the cathode has been properly heated; a safety circuit of the kind usually employed in rectifiers containing such valves has therefore been adopted in the present case.

This circuit is depicted in *fig. 8*, which shows the use of a bimetallic switch to permit the closing









Fig. 9

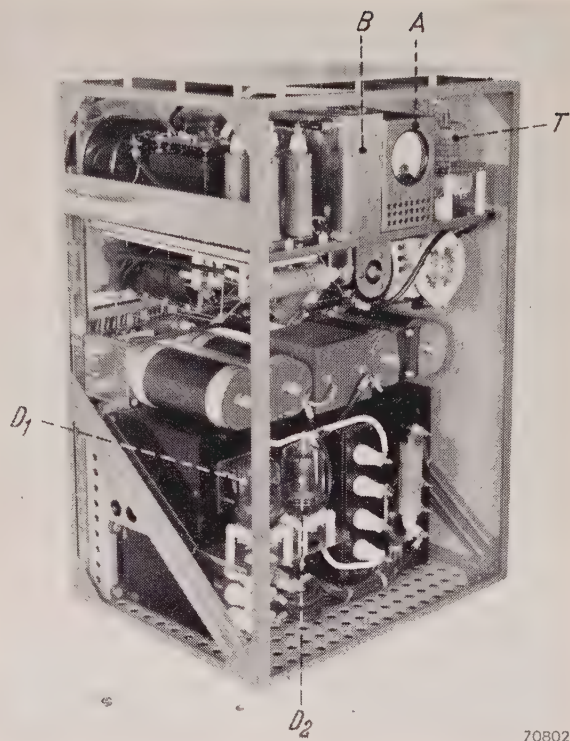


Fig. 10

Fig. 9. Front view of the rectifier. From left to right and from top to bottom: voltmeter for the direct voltage, milliammeter for the D.C., controls for the direct voltage (rough, vernier, continuous), pilot lamp, push buttons operating  $S_9$  and  $S_{13}$ - $S_{14}$  (fig. 8), second pilot lamp, fuses on A.C. side (one is reversible for 110 and 220 V), H.T. cable, mains lead. Consumption on full load 700 W. Dimensions: base 445 mm  $\times$  350 mm; height 790 mm. Weight 80 kg.

Fig. 10. Interior of the rectifier (side and back panels removed).  $A$  = amplifier (with milliammeter in anode circuit of the EF 42).  $B$  = battery box.  $T$  = control valve TB 2.5 300.  $D_1$ ,  $D_2$  = diodes.

*The transient state immediately after the generator is switched on*

When the H.T. transformer is switched on, the amplifier operates for a time under abnormal conditions. The output voltage  $V_0$  rises gradually from zero and, until it attains the value at which the stabiliser tubes ignite, these tubes pass no current and no anode current flows in valve II

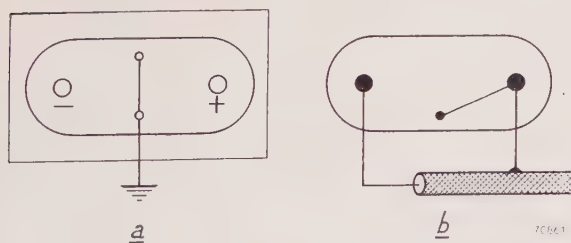


Fig. 11. a) The direct voltage amplifier sockets. b) Corresponding plug on the H.T. cable. In one position of the plug the negative pole and outer conductor of the cable are earthed, with the positive pole to inner core; with the plug reversed, the positive pole and outer conductor are earthed and the negative pole is connected to the core.

(fig. 5) and neither — as the screen-grid voltage is zero — in valve III.

This abnormal condition entails some risk that a state of equilibrium may be reached in which the value of  $V_0$  is not by any means the desired one. The manner in which such a spurious state of equilibrium may arise is demonstrated in fig. 12, in which curve I was plotted using a variable direct voltage  $U_g$  from a separate source<sup>7)</sup> as grid bias for the control valve, instead of the potential across  $R_5$  (fig. 5). A fixed load resistance  $R_0$  was connected across the output terminals. The input voltage  $V_i$  was maintained at a constant value by adjusting the primary alternating voltage. Curve I, fig. 12, represents  $U_g$  as a function of  $V_0$ . For a value of 900 V for  $V_0$  the curve shows that, with the given values of  $R_0$  and  $V_i$ ,  $U_g = -59$  V (point P). If  $U_g$  be now replaced by the potential across  $R_5$ , the question arises as to the manner in which the voltage  $U_{R_5}$  will vary immediately

<sup>7)</sup>  $U$  is assumed to represent the potentials with respect to the positive line,  $V$  those with respect to the negative line.



after the generator is switched on, i.e. when  $V_0$  rises from zero. If this takes place in accordance with curve 2, which does not intersect 1 before the point  $P$ , equilibrium will occur at the desired value

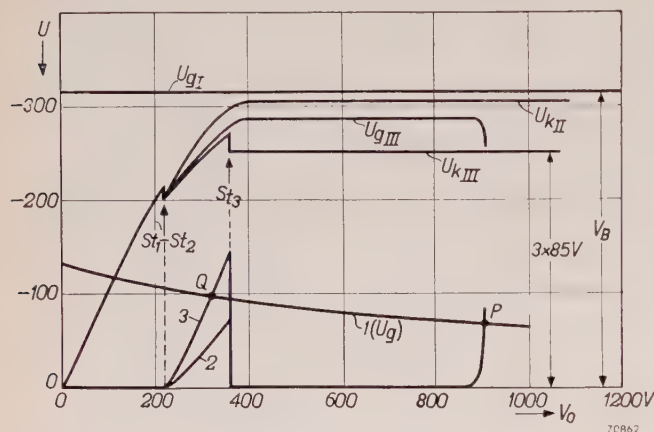


Fig. 12. Curve 1: grid bias  $U_g$  required for the control valve at  $V_i = 3000$  V and load resistor  $R_0 = 18\,000\ \Omega$  to yield the values of  $V_0$  plotted as abscissae. Curves 2 and 3: voltage  $U_{R5}$  across resistor  $R_5$  (fig. 5) as a function of  $V_0$ . Curve 2 intersects 1 only at point  $P$ , and  $V_0$  reaches the required value of 900 V. Curve 3 intersects 1 at a previous point  $Q$ , and  $V_0$  finds a state of equilibrium at too low a value.

The other curves represent the potentials  $U_{kII}$  and  $U_{kIII}$ ,  $U_{gI}$  ( $= -V_B$ ) and  $U_{gIII}$  (see fig. 13) as functions of  $V_0$ . The arrows indicate the values of  $V_0$  at which the tubes  $St_1$ ,  $St_2$ ,  $St_3$  ignite.

of  $V_0$  (900 V). But if  $U_{R5}$  should develop in accordance with curve 3, which cuts curve 1 not only at  $P$ , but at a previous point  $Q$ , equilibrium will be established at  $Q$ , that is to say,  $V_0$  will assume a value that is much too low. It is therefore important to know  $U_{R5}$  as a function of  $V_0$  from  $V_0 = 0$  onwards.

That this function will assume the shape of curve 2 or 3 in fig. 12 is at once apparent from fig. 5. So long as  $V_0$  does not reach the value at which the series-connected stabiliser tubes  $St_1$  and  $St_2$  ignite — which takes place at about 210 V — the pentode, being without screen voltage, passes no current, and  $U_{R5}$  is zero. From the moment that  $St_1$  and  $St_2$  ignite, the pentode passes both screen and anode current and the latter also flows in valve  $II$ . The anode current of the pentode depends on the magnitude of the voltage between the grid and the cathode, that is, the difference between the potential  $U_{gIII}$  of the control grid and the potential  $U_{kIII}$  of the cathode. These potentials, both with respect to the positive line, are also shown as functions of  $V_0$  in fig. 12. When  $V_0$  attains the value at which the third stabiliser tube ignites, the voltage across this tube drops suddenly from the ignition voltage (100–125 V) to the working voltage (85 V), and the difference between  $U_{gIII}$  and  $U_{kIII}$

is increased by 15 to 40 V, that is, to the point where the pentode is cut off and  $U_{R5}$  again drops to zero.

What we now have to guard against is that the curve  $U_{R5} = f(V_0)$  does not prematurely intersect curve  $U_g = f(V_0)$  (fig. 12), which means that the voltage  $U_{R5}$  for igniting  $St_3$ , and hence also the anode current of the pentode, must be kept sufficiently low, i.e. that the difference between  $U_{gIII}$  and  $U_{kIII}$  must remain sufficiently negative.

The resistor  $R_{10}$  (fig. 5 and fig. 13) is the means of affecting  $U_{kIII}$  in such a way as to ensure the desired conditions. From fig. 13 it will be seen, assuming  $St_1$ ,  $St_2$  to have ignited (their working voltage being 85 V), but not  $St_3$ , that:

$$-U_{kIII} = 170 - \frac{R_{10}}{R_6 + R_{10}} (V_0 - 170).$$

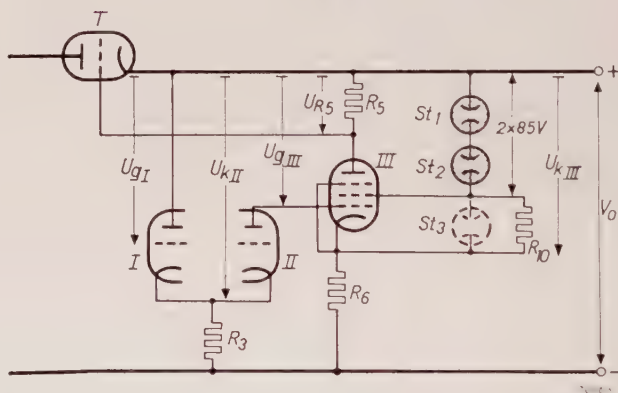


Fig. 13. In this diagram it is assumed that  $V_0$  has risen to the value where  $St_1$  and  $St_2$  have ignited, but not  $St_3$ . Potentials  $U_{gI}$ ,  $U_{kII}$ ,  $U_{gIII}$ ,  $U_{kIII}$  and  $U_{R5}$  are all negative with respect to the positive line. These are shown as functions of  $V_0$  in fig. 12. Other references as in fig. 5.

A reduction in the value of  $R_{10}$  therefore results in a drop in  $-U_{kIII}$ , i.e. an increase in the potential of the cathode of valve  $III$ . By making  $R_{10}$  small enough we can ensure that this potential will be maintained at a value that is sufficiently higher than that of the control grid.

A lower limit is set for  $R_{10}$  by the condition that in the stable state the current passed by the tube  $St_3$  must not be too small. For efficient stabilisation a current of at least 1 mA should flow in this tube. With a further 1 mA cathode current in the pentode, there remains of the current flowing in  $R_6$  (which is 4.45 mA in the stable state) at most 2.45 mA for  $R_{10}$ . It thus follows that with a working voltage of 85 V for  $St_3$ :

$$R_{10} \geq \frac{85}{2.45 \times 10^{-3}} = 35\,000\ \Omega.$$

Once this condition is satisfied, another risk entailed by a reduction in  $R_{10}$  is removed as well, viz. that when  $V_0$  rises, the voltage across  $R_{10}$  might not be sufficiently high to ignite



$St_3$ . It is found that the minimum value of  $R_{10}$  necessary for infallible ignition of  $St_3$  is 23 000  $\Omega$ .

The highest value that can be attributed to  $R_{10}$  without resulting in a premature flow of current in the pentode is 95 000  $\Omega$ . The actual value of  $R_{10}$  has thus been placed at 50 000  $\Omega$ .

Regarding the form of the curves  $U_{gI}$ ,  $U_{kII}$ ,  $U_{gIII}$  and  $U_{kIII}$  in fig. 12, the following may be added.

The potential  $U_{gI}$  of the grid of valve *I* (fig. 5) with respect to positive is  $-V_B$ . Current can flow in this valve only when the cathode potential  $U_{kII}$  approximates to the grid potential  $U_{gI}$ . As long as  $St_1$  and  $St_2$  have not ignited and valve *II* and resistor  $R_3$  are thus not passing current,  $U_{kII} = -V_0$  ( $U_{kIII}$  and  $U_{gIII}$  are then also  $-V_0$ , see fig. 12). When  $St_1$  and  $St_2$  ignite at  $V_0 \approx 210$  V,  $U_{kII}$  is still so far removed from  $U_{gI}$  that *I* is still not conductive. But, once  $St_1$  and  $St_2$  have ignited, current commences to flow in *II* and  $R_3$ . Because of the voltage drop across  $R_3$ ,  $U_{kII}$  now rises less quickly than  $V_0$ , and it is only when  $V_0$  is in the region of 400 V that *I* commences to work. From that point onwards  $U_{kII}$  remains constant at  $\approx -V_B$ .

In these circumstances the difference between the anode and cathode potentials of *II* does not exceed about 15 V and curve  $U_{gIII}$  (= anode voltage of *II*) therefore runs parallel with and close to  $U_{kII}$ .

When  $St_3$  has ignited,  $-U_{kIII}$  is constant at a value equal to the sum of the three working voltages, viz. 255 V.

**Summary.** A rectifier has been designed for laboratory work in the field of decimetric and centimetric waves, to provide a stabilised voltage that will ensure a very constant frequency from transmitting valves fed from the rectifier. Relative fluctuations in the output voltage are only 1/3000th of those in the mains supply. The internal resistance is less than 20  $\Omega$  for H.F. currents. With a direct voltage output of 5000 V the ripple and noise do not exceed 10 mV.

This result is attained by employing a well-known method based on the use of a triode control valve which functions as variable resistor in the D.C. circuit. A fraction of the output voltage is compared with a reference voltage and the difference is amplified in a direct voltage amplifier of which the output voltage is used as grid bias for the control valve. The circuit of this amplifier, which is fed from the high output voltage from the rectifier, is of an unusual type. It comprises two stages giving an overall gain of 3000, with an input stage giving no gain (cathode follower). The reference voltage is supplied by dry batteries (300 V). The output voltage is variable from 900 to 5100 V in four stages of 1200 V, 6 stages of 200 and continuously over a range of slightly more than 200 V. The high direct voltage and gain entail special conditions as regards assembly and quality of the various components.

Some of the precautionary measures taken with a view to safeguarding the equipment and preventing instability are discussed, as is also the method of preventing the output voltage from assuming too low a level of equilibrium.



## APPARATUS FOR MEASURING LIGHT DISTRIBUTIONS

535.245.002.54



The apparatus shown in the reproduction was designed in the Light Technical Laboratory at Eindhoven for measuring the luminous intensity, in all directions, of light sources of all kinds and sizes. The intensity is measured by means of a photo-electric cell, which is set up on the right-hand

side at a large distance, in a fixed position; in the drawing on the opposite page the photo-electric cell is situated on the extended part of the dotted line *A*.

The apparatus consists mainly of a sort of cage which can be rotated about the above-mentioned



horizontal dotted line as its axis. In the cage a large flat mirror (*S*) is mounted at an angle of 45° to this axis. The mirror always reflects a beam of light from the light source to be examined,

cage is turned. Therefore the photo-electric cell “scans” the emitted light in a meridian round the light source at each complete revolution of the cage.

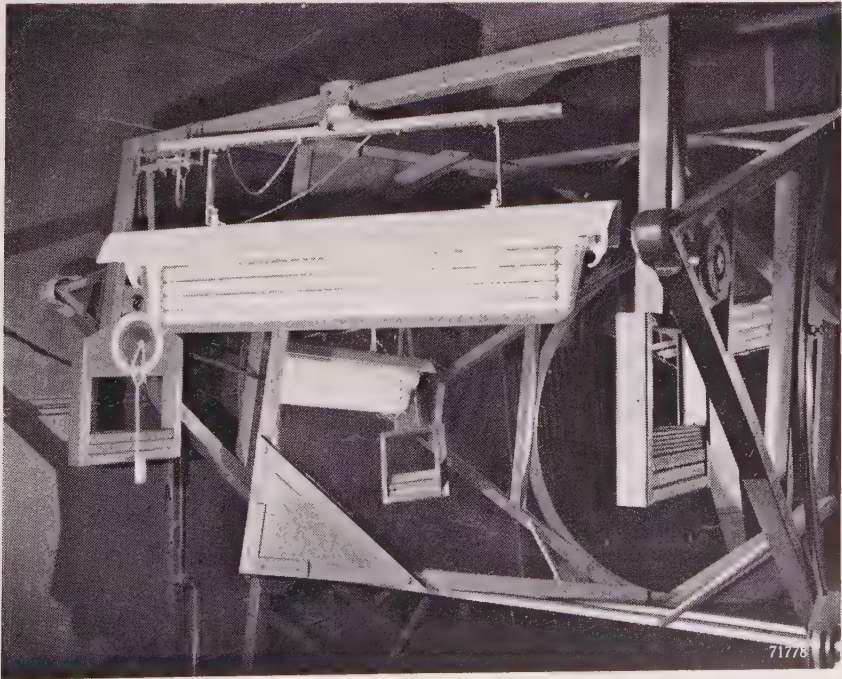
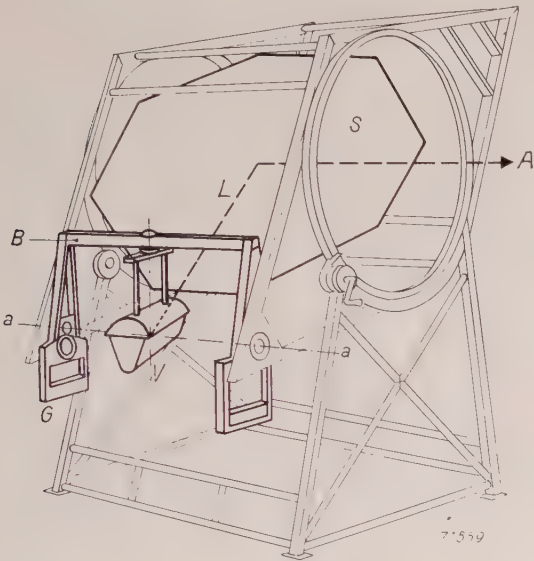
By turning the light source through a small angle about the (always vertical!) axis *V* after each complete revolution, a new meridian can be scanned each time, and in this way the total light distribution can be determined.

The close-up below shows the suspension of the light source in the bracket in a different position. By turning the bracket upside down, the light distribution of standing fittings can also be measured.

Of course the construction would have been simpler if the light source itself had been placed with its centre on the axis *A* and was revolved about two axes perpendicular to *A*. However, this would involve tilting over the light source from its normal burning position; several types of lamps cannot be used in any position; with other types it would lead to inaccurate measuring results. For these reasons the apparatus would not be suitable for these lamps. Therefore, in other apparatus designed for the same purpose, the light source is set up in a fixed position, either the photo-electric cell itself or one or more mirrors, which lead the light from the source to a fixed photocell, swinging about the light source. The depicted apparatus compares favourably to such constructions in that, with relatively small dimensions of the apparatus, the photo-electric cell can be placed at a very large distance <sup>1)</sup>

emitted in the direction of the axis *L*, in the direction *A* to the photo-electric cell, independent of the position of the cage. The light source — in the photographs it is a fitting for street lighting with four fluorescent lamps (“TL” lamps) — is suspended in a bracket (*B*), which can turn freely about the secondary axis *a-a* passing through the centre of the light source. By means of weights (*G*) the bracket, and consequently the light source, are always kept in a vertical position when the

<sup>1)</sup> This is necessary for light sources with a large “photo-metrical boundary distance” (e.g. projectors); see Philips techn. Rev. 9, 114, 1947.





and that no excentric movement of a large, heavy mirror is necessary. In our case the acting forces and couples are mainly symmetrical and relatively small, so that inadmissible inaccuracies in the setting of the measuring direction due to all sorts of deformation are avoided with a relatively light and simple construction.

The last-mentioned point of view has also lead to the peculiar cage construction. Due to the axially emitted beam of light, the rotating construction on this side could not be supported by a normal bearing. With one single bearing on the other side the required rigidity could only have been obtained at the cost of a very heavy and expensive construction. The solution shown in the figures with two large rings each resting on two roller bearings in a pedestal is light and yet rigid. As a consequence

of the bearings being fitted on both sides, we only observe vertical sagging at the suspension points of the fitting and mirror and no angular deformations, which are much more objectionable if accuracy of measurement is required. The sagging of the bracket where the fitting is suspended is less than 2 mm with a fitting weighing 100 kg, and, moreover, of no consequence, as it is automatically taken into account during the adjustment of the fitting. The largest observed sagging of the mirror is less than 0.1 mm. Consequently the angle measurements are most reliable: in every meridian the angle ("elevation") can be adjusted to  $1/20^\circ$ , the angle of the meridian itself ("azimuth") to  $1/12^\circ$ . Accuracies of this order are required, among others, for some lights for car and airfield lighting.

L. de WIT.



## HARDENING OF METALS

by J. L. MEIJERING.

539.53:621.785.61

*For many technical applications a metal having a high degree of hardness is required. This hardness is obtained by alloying metals which are not hard enough in their pure condition with other elements: sometimes some sort of heat treatment or a deformation process follows. The most important hardening methods are quenching (steel hardening) and precipitation hardening. A new method was developed in the Philips Laboratory at Eindhoven, viz. the so-called oxidation hardening. This method, which is still in the experimental stage, is discussed briefly in this paper. We have availed ourselves of the opportunity to discourse at some length on the hardness of metals in general and on the physical phenomena which form the basis of the hardening process in particular.*

## Hardness

The hardness of a material in the widest sense of the word is understood to be the resistance which must be overcome to deform this material plastically, i.e. permanently. The resistance against elastic deformation, i.e. deformation which returns automatically to zero after removal of the deforming forces, is governed by the elastic properties of the material. We shall not discuss this subject here.

In a narrower sense hardness may be defined as the resistance against impression. This hardness is often expressed in so-called Vickers units. In quite a number of cases "hard" may also mean: "of great tensile strength". Generally speaking, a fair correlation is found between tensile strength and Vickers hardness for a given material.

The crystal structure is of importance with regard to the plastic properties: as the number of suitable glide planes in the crystal is larger, so the malleability will be greater. Thus the great hardness and brittleness of silicon and germanium is attributable to the special structure of these elements, viz. the so-called diamond structure, in which each atom has four adjacent neighbours (fig. 1d) and in which no proper glide planes occur. The occurrence of the diamond structure (the prototype of which, viz. diamond, is extremely hard) in silicon and germanium is closely connected with the homopolar character of these "semi-metals".

These semi-metals will not be discussed here; we shall turn our attention to the body-centred cubic

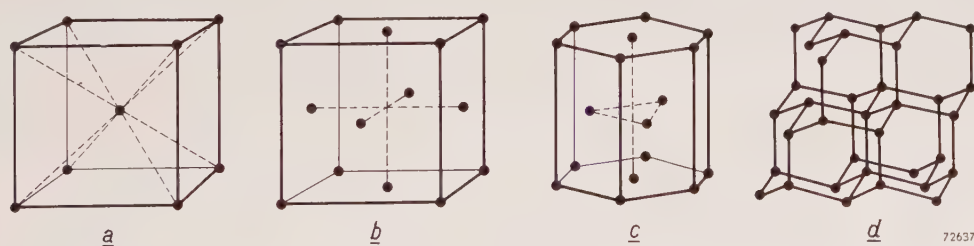


Fig. 1. The three structures in which most metals crystallise: a) body-centred cubic structure, b) face-centred cubic structure, c) hexagonal close-packed structure. The diamond structure is also depicted (d). The atoms have been drawn relatively much too small in these figures; they should touch each other.

In plastic deformation layers of atoms move along "glide planes" in the crystal. These glide planes are often those layers of atoms for which the distance between two successive parallel lattice planes is a maximum, and in which the packing in the plane (layer of atoms) itself is densest. Fig. 1 depicts a few crystal structures of frequent occurrence. In face-centred cubic crystallised metals e.g. (fig. 1b) the planes of closest packing are the (111)-planes.

arrangement (fig. 1a), to the face-centred cubic and to the hexagonal close-packed metals. The majority of the real metals has one (or more, in elements showing allotropy) of these three structures<sup>1)</sup>.

<sup>1)</sup> We wish to observe that a true metal such as manganese is also very hard and brittle if it occurs in the  $\alpha$  or  $\beta$  modification, since the crystal structures of  $\alpha$  and  $\beta$  manganese are complicated and have no well-defined glide planes. On the other hand,  $\gamma$  manganese, which is stable at high temperatures has a face-centred structure and is ductile.

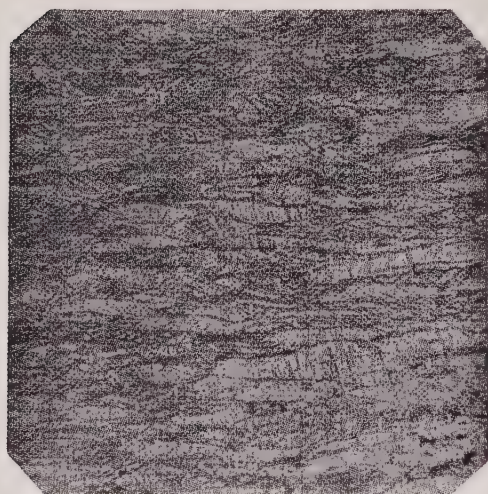


The hardness of a metal in its pure condition is often not sufficient for technical purposes. In the majority of cases attempts are made to increase the hardness in one way or other, although for some applications it may be desirable that the metal is not hard. Generally speaking it is not possible to decrease the hardness permanently; on the other hand, various methods exist for increasing the hardness, which under the circumstances is a great advantage.

### Hardness of pure metals

On comparing various pure metals of the same crystal structure, the hardness will (at a given temperature) be greater as the melting point of the

The hardness of one certain (pure) metal at a given temperature may be varied considerably, as mentioned before, chiefly in an increasing sense. The hardness increases by plastic deformation (rolling, drawing, hammering): this work-hardening is based on the fact that the lattice is distorted by the gliding of the atoms along the glide planes<sup>2)</sup>. It stands to reason that this method of hardening by rolling etc. can be realised with ductile metals only. The work-hardened material can be resoftened by heating: if the temperature does not rise too high and the material is not too strongly deformed, recovery may occur and the lattice distortions rearrange themselves in such a way that the total distortion is diminished. On heating to a higher tem-



a



b

Fig. 2. Recrystallisation of brass with 20% zinc; a) cold, rolled, b) subsequently annealed (taken from P. Goerens, Einführung in die Metallographie, Wilhelm Knapp, Halle 1932).

metal is higher. This is illustrated by the series Pb, Al, Cu, Ni, Pt. This is not surprising: a lower melting point means a weaker cohesion between the atoms. The hardness of one and the same metal decreases with rising temperature: the stronger thermal agitation involves an easier surmounting of the cohesive forces in the gliding process.

We wish to observe, in order to avoid misapprehension, that although greater hardness is, broadly speaking, often accompanied by greater brittleness (smaller elongation), this need not always be the case, even with one and the same metal. With the above-mentioned face-centred cubic metals, both hardness and elongation increase with decreasing temperature.

perature or after stronger deformation, recrystallisation occurs: a number of nuclei grow to new, undistorted crystals with orientations differing from those of the deformed old crystals which they consume (fig. 2). Consequently, the hardness decreases both on recovery and with recrystallisation. If the plastic deformation takes place at a temperature higher than the "recrystallisation temperature" (which naturally varies widely for different

<sup>2)</sup> We may refer the reader to J. D. Fast, Strain ageing in iron and steel, Philips techn. Rev. 14, 60-67, 1952 (No. 2), for a detailed picture of these irregularities (dislocations).



metals, and also for a given metal depends greatly on the specific deformation process and the annealing treatment) no appreciable work-hardening will occur (hot rolling). With a metal having a low melting point the recrystallisation temperature may even lie below room temperature (as with lead).

The hardness of a completely recrystallised ("soft annealed") metal still depends more or less on the grain size. A grain boundary means a change in the orientation of the lattice and consequently in the direction of the glide planes, and therefore interferes with the deformation process. For this reason, very fine crystalline copper e.g. will be harder than a Cu single crystal.

At high temperatures, where creep occurs, i.e. a deformation that increases with time when a constant load is applied, the reverse is observed: a Cu single crystal is found to creep much slower than polycrystalline copper. Similarly, well annealed copper does not creep as quickly as copper that is less completely recrystallised after deformation. Apparently at these high temperatures the lattice distortions promote the deformation process. We may attempt to explain this by suggesting that the disappearance of all types of lattice faults, while heating under stress, causes creep. Experiments in this field have not yet given a decisive answer to this phenomenon.

A very much distorted lattice of a metal may also be obtained by building up the lattice in circumstances (e.g. at a rather low temperature) in which many lattice faults will occur. Very hard nickel<sup>3</sup>) may, for example, be obtained by electrolytical deposition under special conditions. The presence of chemical impurities, in very small concentrations, chiefly dissolved gasses, e.g. hydrogen, is probably also of importance.

### Hardness of alloys

For simplicity's sake we shall confine ourselves in the following to binary alloys. Alloys with three or more components show no essentially new points.

There are two main types of solid solutions: substitutional solid solutions, e.g. Fe-Ni, and interstitial solid solutions, e.g. Fe-C. In the first case, with which we are nearly always concerned if both components are metals, Fe atoms are displaced by Ni atoms on, in most cases irregularly distributed, lattice points. In the second case — which occurs if the added element is a non-metal having atoms which are small in comparison to the atoms of the parent metal — the foreign atoms are built in at

places (interstices) between the lattice sites of the parent lattice.

We may expect the hardness of the solid solution to be greater than that of the parent metal (the latter must be understood to mean the metal with the greatest concentration), since interstitial atoms will disturb the order of the lattice, especially if they are larger than the normal interstices in the lattice of the pure parent metal. Substituted atoms will also deform the lattice, proportionately more as the atoms of the metal added to the alloy differ more in size from those of the parent metal. In a dilute solid solution the hardness increase is approximately proportional to the concentration in foreign atoms, and according to a rule of Norbury, when various metals are added, the change in hardness per atomic per cent addition in a given parent metal (e.g. copper) is approximately proportional to the absolute value of the difference in atomic diameters, as determined from the lattice constants of the pure metal<sup>4</sup>).

It is only on approximation that a metal lattice is composed solely of (e.g. spheric) atoms. The valency electrons are more or less detached from the ions on the lattice points and together form the "electron gas", which is responsible for a high electrical and thermal conductivity. For this reason Norbury's rule should not be taken too literally, especially if the differences in atomic diameter are small. Silver and gold e.g. have practically the same atomic diameter (the same face-centred cubic structure with a difference in lattice constant of only  $2 \frac{0}{100}$ ), but still silver is appreciably hardened by alloying it with 5 at. % gold (and vice versa). It is, however, a fact that the hardness increases more quickly with the concentration if the differences in atomic diameter are large than if they are small.

Viewed from this angle, the question: "Given a certain metal, in what manner can it be hardened as much as possible by alloying it with e.g. 5 at.% of another element?" should be answered as follows: "Add an element the atoms of which differ as much as possible in size from the parent metal, e.g. alloy copper with lead". One circumstance, however, prevents us from acting on this principle. For the substitution of very large (or very small) atoms in a metal lattice costs so much energy, that the solubility becomes very small and the greater part of

<sup>3</sup>) The same applies to silver, see E. Raub, *Z. Metallk.* **33**, 87, 1947, and to chromium.

<sup>4</sup>) A. L. Norbury, *Trans. Faraday Soc.* **19**, 586, 1924. M. A. Meyer (Diss. Delft 1951) also found this for solid solutions of nickel. Other authors found a hardening proportional to the square of the "misfit", for Cu and Ag solid solutions (see Diss. Meyer, page 57); Mott and Nabarro ("Strength of solids", Bristol Report 1948) also found this theoretically.



the metal added to the alloy occurs in the form of a second phase rich in this metal; the homogeneous alloy changes into a heterogeneous one.

### Heterogeneous alloys

The hardness of an alloy consisting of two phases is, roughly speaking, a volume average of the hardnesses of the two phases, measured as micro hardnesses *in situ*<sup>5)</sup>. In most cases, and nearly always

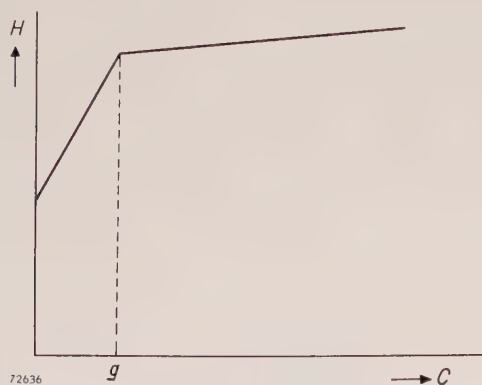


Fig. 3. The hardness  $H$  of a binary alloy increases as the concentration  $C$  of the alloy component in the parent metal increases, and that approximately proportional to this concentration, as long as the alloy component can still be dissolved in the parent metal (homogeneous alloy). If the saturation point  $g$  is passed, however, and the second phase is being formed, the alloy becoming heterogeneous, the change in hardness chiefly depends on the hardness of the new phase. The increase in hardness becomes much slower and sometimes even changes into a decrease.

if the concentration of the element added to the alloy is relatively small, the total hardness develops as shown in fig. 3. Consequently, the hardness is smaller in the heterogeneous field after passing the saturation point than it would be if the solution was supersaturated homogeneous (as shown by extrapolation); the latter hardness can in some cases be experimentally determined in supersaturated solutions. If the atomic diameters of both elements differ more than 15 to 20 %, the saturation point ( $g$  in fig. 3) is always very low, even at high temperatures (Hume-Rothery)<sup>6)</sup>. In these cases, according to Norbury, the hardness will rise steeply in the homogeneous field, but the attained increase in hardness at the saturation point is still negligible because the homogeneous field is so narrow. After passing the saturation point, the hardness rises less

steeply, or in some cases even falls, as e.g. in the system Cu-Pb, where the second phase, nearly pure lead, is softer than the parent phase.

### Hardening methods

It is possible to bring large atoms (or molecules) into the lattice and so harden the metal, albeit that the resulting alloy is thermodynamically not stable<sup>7)</sup>. This is very important for practical hardening methods.

The chief hardening methods are those applicable after the forming, e.g. rolling or drawing of the metals, and which result in a greater hardness of the material than it possessed during the forming operation. We may distinguish three methods: quenching from the solid phase (martensitic hardening), precipitation hardening and oxidation hardening, which will be discussed in succession. The first two methods have been known for a long time already, we shall therefore only discuss them briefly. The third method, the oxidation hardening, was developed in the Philips Laboratory at Eindhoven during the last decade: this method will be discussed somewhat more extensively in this paper.

A relatively hard metal may also be obtained by a special treatment before or during the forming operation. One of these methods is quenching liquid alloys<sup>8)</sup>. In molten copper as much as 10 at. % of lead is soluble, and it is conceivable that, provided the cooling is effected quickly enough, a considerably larger concentration of lead in the solid solution may be obtained than the 0.001 to 0.01% soluble in copper in the solid phase. The desired result might also be obtained both by quick condensation from the vapour phase of two metals at the same time, and by simultaneous electrolytical deposition.

There are not many data on the two first-mentioned methods. The electrolytical method is more important for practical applications, with this understanding however that very great hardnesses are also obtainable by electrolytical deposition in a special manner of e.g. nickel alone (see above)<sup>9)</sup>.

One of the processes by which the hardness of a metal is altered during the forming operation is the work-hardening already mentioned above. We shall not enlarge on this subject here either, but we do wish to point out in passing the peculiarities that may occur during the work-hardening of alloys. A most remarkable case occurs in the system aluminium-zinc. Zinc with approximately 20% aluminium can in some circumstances be much softer than zinc and aluminium themselves<sup>10)</sup>.

<sup>5)</sup> Roughly, because analogous to the influence of the crystal size in homogeneous alloys, the hardness of a heterogeneous alloy will be greater as the second phase is more finely divided.

<sup>6)</sup> Although a small difference in atomic diameter is required for a proper solubility, it does not guarantee this. The solubility of Fe in solid Cu for example, is only very small, although Fe and Cu differ only 1% in atom diameter (and intermetallic compounds are not formed).

<sup>7)</sup> Cold-worked metals and alloys do not possess this quality either, nor strictly speaking do poly-crystalline metals and alloys (surface tension!).

<sup>8)</sup> For this subject see e.g. H. Bückle, *Z. Metallk.* **43**, 82 1952.

<sup>9)</sup> Raub and Engel (*Z. Metallk.* **41**, 487, 1950) obtained strongly supersaturated solid solutions of Cu with Pb, and of Ag with Bi electrolytically. A great hardness was indeed obtained.

<sup>10)</sup> A. A. Bochvar and Z. A. Sviderskaya, *Bull. Acad. Sci. U.R.S.S. Classe sci. techn.* 1945, 821; see also *Chem. Abstr.* **41**, 2375, 1947. F. Sauerwald *Arch. Metallk.* **3**, 165-173, 1949, studied this alloy independently; the elongation was found in some cases to amount to 180%.



Until now it is a unique thing for a binary alloy to be softer than its two components. This is a case of non-equilibrium conditions: the adjustment to equilibrium is promoted by the deformation and conversely the deformation is favoured by the fact that the atoms are rearranging themselves.

### Martensitic hardening

In martensitic hardening we try to take advantage from the transition of a solid solution of one solid phase into a phase of a different structure, e.g. of the transition face-centred cubic into body-centred cubic, which may occur when cooling various iron alloys (austenite-ferrite). This method will not prove very successful in the case of substitutional solid solutions: if the atoms differ too much in size, the solubility is small for both structures. With interstitial solid solutions the case is different. In certain circumstances the deformation of the lattice will increase roughly proportionally with  $(r-r_h)/r_h$ <sup>11</sup>), in which  $r$  is the atom radius of the dissolved element and  $r_h$  the radius of the inscribed sphere of an interstice in the (undistorted) parent lattice. It is possible for  $r_h$  to differ considerably for both structures, e.g. it can be larger for the phase stable at high temperature than for the phase stable at low temperature. In the first-mentioned phase the solubility for too large an atom will then be much higher.

On quenching, from e.g. 900 °C, of  $\gamma$ -iron (austenite) with e.g. 4 at. % (0.9% by weight) carbon dissolved in it,  $\alpha$ -iron (ferrite) is obtained in which carbon is found in supersaturated solution (martensite). The strong hardening occurring in this case (i.e. the classical steel hardening that has already been applied for centuries) may be ascribed to the strong deformation caused by C atoms occupying too narrow places in the ferrite lattice.

We may roughly calculate the difference in lattice deformation by an interstitial atom in the body-centred or face-centred cubic structures respectively as follows. Assume the lattice deformation per atom dissolved to be proportional to  $n(\Delta r)^2$ . In this formula  $n$  stands for the number of adjacent neighbours per atom: 12 for a face-centred cubic substitutional solid solution, 6 for a similar interstitial solid solution, 8 for a body-centred cubic substitutional solid solution, 2 or 4 for a similar interstitial solid solution.  $\Delta r$  stands for  $r-R$  in a substitutional solid solution, viz. the difference in atomic radii of the dissolved metal ( $r$ ) and the parent metal ( $R$ ); in an interstitial solid solution  $\Delta r$  stands for the difference between the atomic radius of the foreign atom and the radius  $r_h$  of the interstices in the parent metal. We have assumed the exponent of  $\Delta r$  to be 2, which follows from purely geometrical considerations for  $n(\Delta r)^2$  is proportional to the volume occupied "doubly", if  $\Delta r$  is small, assuming the positions and radii of the atoms present in the metal to remain unchanged

<sup>11</sup>) If  $r > r_h$ , which is normally the case.

when the foreign atom is brought into it. An analysis of the increase in hardness proportional to the concentration of homogeneous alloys (Nörbury's rule) shows that actually the exponent of  $\Delta r$  lies probably between 1 and 2 (compare note <sup>4</sup>)).

Let us consider the face-centred cubic structure. An interstice has only half as many adjacent neighbours as an atom has. Moreover, according to elementary mathematics,  $r_h = R(\sqrt{2}-1)$ . When discussing the heterogeneous alloys we saw that, generally speaking, a difference in atomic radius of 15% is still only just tolerable for a substitutional solution to be formed. According to our formula, a ratio  $r/R = 1 + u$  for substitutional solid solutions corresponds, as regards lattice deformation, to a ratio  $r/R = \sqrt[4]{2-1+u}/2$  for interstitial solid solutions. If we substitute the value 15% for  $u$ , we find that the addition of atoms with a diameter of approximately 63% of the diameter of the atoms of the parent metal would be just tolerable for an interstitial solution to be formed.

In the body-centred cubic structure two different kinds of interstices are suitable for occupation<sup>12</sup>). The largest interstices have four adjacent neighbours, and in this case  $r_h = R(\sqrt[5]{3}-1)$ ; the smaller interstices have two adjacent neighbours, and  $r_h = R(\sqrt[4]{3}-1)$ . In this case we find that  $r/R = 1 + u$  for substitutional solid solutions corresponds to  $r/R = \sqrt[4]{\frac{4}{3}-1+2u}$  for interstitial solution in the smallest interstices, and to  $r/R = \sqrt[5]{\frac{5}{3}-1+u/2}$  for interstitial solution in the larger interstices. Substitution of  $u = 0.15$  gives 0.455  $R$  and 0.50  $R$  respectively for the largest atom diameter tolerable for interstitial solution to be formed.

We shall now consider the steel hardening somewhat closer. The radii of C atoms and N atoms amount to 61% and 55% of the radius of the iron atom respectively. This explains why carbon and nitrogen dissolve rather well in austenite (face-centred structure), and only to a very low degree in ferrite (body-centred structure); a transition of austenite into ferrite on quenching should give a strong deformation and consequently great hardening.

We shall have to bear in mind that the ratio of the atomic radii is not the only factor determining the solubility; however, we shall not discuss the other factors in this paper.

### Precipitation hardening

A metal lattice will also be considerably deformed if we succeed in producing groups of atoms or molecules in the lattice in one way or other. Precipitation hardening is an example of a hardening method based on the first-mentioned process. Precipitation hardening (as also martensitic hardening) is often applied in practice; the following may serve as an example. Aluminium atoms and copper atoms differ  $\pm 10\%$  in atom radius. At room temperature the solubility of copper in aluminium is practically zero, but at 550 °C (at the eutectic temperature) 2 at. % Cu are soluble in aluminium. After quenching from this temperature to room

<sup>12</sup>) See J. D. Fast, Ageing phenomena in iron and steel after rapid cooling, Philips techn. Rev. **13**, 165-171, 1951 (No. 6).

temperature the solid solution is supersaturated. The hardness is normal: somewhat harder than in the stable state, in which practically all copper is precipitated in the form of coarse  $\text{CuAl}_2$  particles as a second phase. This stable final state can never be attained at room temperature, since the diffusion rate is much too low. When annealing at temperatures of roughly  $300^\circ\text{C}$  (where the solubility is still low), the final state is approached much closer within a measurable space of time. It is of importance that this final state is reached via intermediate states, in which the hardness is considerably greater than in the initial and final states; there is thus a maximum hardness. These intermediate states, which are characterised by the presence of knots of Cu-atoms which deform the lattice (see the article referred to in note <sup>12</sup>) can be fixed at room temperature by cooling. By a proper duration and temperature of the annealing treatment, very strongly hardened materials can be obtained in this way. A considerable hardness is particularly obtained by "annealing" at temperatures below approximately  $200^\circ\text{C}$ . In proportion as the temperature is lower, it takes more time to reach the optimum. Hardening also occurs at room temperature, but the maximum is not attained.

We have already observed before in this article that a heterogeneous alloy is harder in proportion as the second phase is more finely divided. In the above-mentioned state of maximum hardness the particles are not visible under the optical microscope; this is a case of transition homogeneous-heterogeneous. The (vague) boundary between homogeneous and heterogeneous states apparently lies at various average particle sizes, judging from various properties.

The existence of such a thing as an optimum size of the particles may be explained by the following purely geometrical reasons, to which no quantitative value should be attributed however. Let  $x$  be the radius of the particles assumed to be spheric, then the concentration of the particles is inversely proportional to  $x^3$ . We have seen that if  $(x-R)$  is small, the deformation of the lattice resulting from the presence of the particles increases with  $(x-R)^p$ , where  $p$  lies between 1 and 2. Assuming this to hold good also for larger values of  $(x-R)$ , the total deformation is proportional to  $(x-R)^{p/x^3}$ , which expression gives a maximum for  $x = 3R/(3-p)$ . According to this reasoning we could expect particles of a few times the atomic radius of the parent metal to be most effective in precipitation hardening. However, the influence of internal stresses and surface tensions renders the quantitative result of the calculation very uncertain.

We wish to observe that in quite a few cases, for example in the system Al-Cu, the particles are not spherical but in the form of platelets. The smallest

dimension, in this case the thickness of the platelets, seems to amount to a few times the atomic radius.

It stands to reason that there will be a tendency towards such submicroscopical precipitation phenomena also in the case of steel hardening. In this case the stable final state consists of practically pure Fe plus coarse carbide particles  $\text{Fe}_3\text{C}$  (or, still more stable, graphite). There is every possibility in this case of an initial "coagulation" resulting in still greater hardnesses than the "ideal" state in which the carbon is divided atomically. Precipitation phenomena will also occur in the above-mentioned electrolytical alloys <sup>13</sup>).

For a more profound discussion of steel hardening and precipitation hardening in general we may refer to the very extensive literature existing on the subject <sup>14</sup>).

### Oxidation hardening

Oxidation hardening, or rather hardening by internal oxidation, is a method of hardening based on the formation of foreign molecules in the crystal lattice of the basic metal.

We wish to discuss this method somewhat more extensively than the methods mentioned earlier in this paper. First of all we shall illustrate the oxidation hardening by an example. On heating in air silver in which 1.5 at. % magnesium has been dissolved (substitutionally), a perceptible quantity of oxygen diffuses into the metal. This oxygen combines with the Mg-atoms to form molecules of  $\text{MgO}$ , and these molecules deform the silver lattice considerably by their large dimensions. Consequently the metal becomes harder; the Vickers hardness increases from about 40 to approximately 170.

The above shows that oxidation hardening is caused by internal oxidation of a component of an alloy. The occurrence of internal oxidation has been known for a much longer time than the hardening phenomena that may accompany this oxidation <sup>15</sup>). Internal oxidation will not in all cases proceed in the manner discussed above. One of the conditions for a process to occur in analogy with that in the example given above is that the oxidising component occurs in a relatively small concentration in the solid

<sup>13</sup>) Raub and Engel <sup>9</sup>) indeed found a hardness optimum, when tempering their electrolytically deposited alloys, and the same is found when tempering martensite. (see K. H. Jack, J. Iron Steel Inst. **169**, 26, 1951).

<sup>14</sup>) We may refer the reader for a survey of the most important phenomena in steel hardening to F. Seitz, The Physics of Metals, chapt. XIV, Mc. Graw Hill, New York 1943. A survey of precipitation hardening is found in G. C. Smith, Progress in Metall Physics **1**, 163-234, 1949.

<sup>15</sup>) See e.g. C. S. Smith, Min. and Met. **11**, 213, 1930; **13**, 481, 1932; K. W. Fröhlich, Z Metallk. **28**, 368, 1936; F. N. Rhines Met. Technol., **7**, T.P. 1162, 1940.



solution, the parent material itself remaining unaffected, so that the metallic properties of the alloy are not lost.

We shall in the following discuss (dilute) binary alloys only. The conditions for the formation of an oxide of the dissolved metal in the metallic parent metal are:

- 1) The oxygen should diffuse more rapidly into the metal than the atoms of the dissolved metal diffuse to the outside.
- 2) The affinity of the dissolved metal for oxygen should be greater than the affinity of the parent metal for oxygen.

As to the first condition: oxygen diffuses rapidly through silver, copper, nickel,  $\gamma$ -iron, titanium and zirconium. The diffusion coefficients  $D_M$  of foreign metals dissolved in one of these parent metals are mutually roughly equal and much smaller than the diffusion coefficient  $D_O$  of oxygen (at the same temperature). It is practically impossible to comply with the second condition for internal oxidation in the case of titanium and zirconium. Consequently, the parent metals suitable for internal oxidation are silver, copper, nickel and iron.

The oxidation in the metal progresses from the outside to the inside. The penetration depth  $d$  consequently increases with time; a good approximation of the penetration depth as a function of time is given by:

$$d = a\sqrt{t}.$$

This can also be deduced theoretically<sup>16</sup>). Fig. 4 gives an illustration of the penetration. The factor  $a$  contains the diffusion coefficients  $D_M$  and  $D_O$  of the



Fig. 4. Internally oxidized silver strip with 1.2 at. % manganese. The strip was heated for 50 minutes in air at 800 °C and afterwards sectioned, polished and etched. The oxidation proceeding from the outside to the inside may be clearly seen: a "rim" covering about  $\frac{1}{4}$  of the thickness of the strip is oxidized, the interior is untransformed.

dissolved metal and the oxygen in the parent metal respectively, and the concentrations  $C_M$  and  $C_O$  in which the dissolved metal and the oxygen occur in the metal (the latter concentration as measured at the metal surface). If the ratio  $C_M D_M / C_O D_O$  is much smaller than unity (e.g. approximately 0.01),

a case which is practically important,  $d$  does not depend on  $D_M$ .

The concentration of the oxide molecules in the oxidized layer (for the present we can neglect the diffusion of these molecules) is, if  $M$  is a bivalent metal, given by

$$C_{ox} = C_M(1 + C_M D_M / C_O D_O),$$

and, consequently, is somewhat higher than  $C_M$  in the above-mentioned case. Naturally, the concentration cannot have this value everywhere in a completely internally oxidized metal, since it is im-

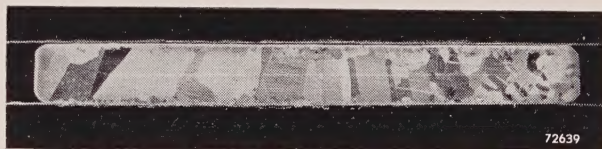


Fig. 5. In a strip completely oxidized internally (in this case silver with 1.5 at. % manganese) a "middle-line" poor in oxidized metal is left.

possible for the number of oxide molecules formed of the dissolved metal to be larger than the number of atoms of this metal. Therefore, if a thin metal strip is completely internally oxidized from both sides, we can expect a very thin zone in the middle of the metal where the oxide concentration is very much poorer. As a matter of fact such a "middle-line" is indeed found, as may be demonstrated by etching (fig. 5). Even after hundreds of hours of annealing the internally oxidized metal, this middle-line will still be found; this proves that the diffusion of oxide molecules is a very slow process indeed. Still it cannot be absolutely neglected, for the oxide molecules will dissociate to a certain extent into metal atoms and oxygen atoms, which do diffuse each by themselves through the metal and can recombine somewhere else in it. In this way coagulation of the oxide molecules to coarser particles can occur indirectly. The smaller the difference in affinity for oxygen between dissolved metal and parent metal, the easier will this take place. Since a coarse precipitate gives no proper hardening (see previous section), the terms of the second condition should be more strict in order to render oxidation hardening possible: the difference in affinity should be large (a quantitative thermodynamic discussion is found in<sup>16</sup>)).

We wish to point out here the difference between internal oxidation of silver alloys and internal oxidation of alloys of a less precious metal, such as copper. In the first case no oxide layer of the parent metal (silver) is formed at the surface of the metal. The oxygen concentration at the surface,  $C_O$ , is

<sup>16</sup>) J. L. Meijering and M. J. Druyvesteyn, Hardening of metals by internal oxidation, Philips Res. Rep. 2, 81-102, 1947 and 2, 260-280, 1947.



then directly proportional to the root of the oxygen pressure in the ambient atmosphere. The less precious metals are covered with an oxide layer (e.g.  $\text{Cu}_2\text{O}$ ) at their surface on annealing in air: the internally oxidized metal only begins under this completely oxidized surface layer. In practice the resulting metal loss during the hardening process can be reduced by replacing the oxygen in the ambient atmosphere by nitrogen after a short time. The oxide already formed will supply sufficient oxygen for the completion of the internal oxidation within the metal.

We shall now give a few examples of the increase in hardness obtained by internal oxidation. By heating the silver alloy with 1.2 at. % magnesium in air for 2 hours at 800 °C, its Vickers hardness rose from the original value 40 to 170. Replacement of the magnesium by aluminium, beryllium and manganese gave final hardnesses of 160, 135 and 140 respectively (with practically the same concentrations of the alloys and equal annealing time). Zinc, tin and cadmium as components of the alloys caused no or only very little hardening. This will be readily understood: the oxides of these elements are not particularly stable.

Alloys of the composition  $(100-x)\% \text{ Cu } x\% M$ , in which  $M$  stands for resp. Be, Al, Mg, Ti or Zr, and  $x$  lies around 1.5, showed an increase in hardness of appr. 40 to 165, 105, 105, 125 and 100 respectively. We found that manganese did not cause considerable oxidation hardening in copper, but it did in silver. This is due to the fact that the affinity of copper for oxygen is greater than that of silver for oxygen (more stringent condition 2). In this connection we wish to observe that it was only possible to bring about oxidation hardening in nickel ( $\text{NiO}$  is still more stable than  $\text{Cu}_2\text{O}$ ), when aluminium was added. Oxidation hardening has not been found in iron alloys.

As said before, as in precipitation hardening, we shall also have to expect coagulation of oxide molecules occurring during the ageing of the alloy to affect the hardness, in the case of alloys hardened by internal oxidation. There may again be an optimum size of the particles giving the greatest hardness. We have not been able to verify this conclusively by experiments. It is beyond the scope of this article to discuss the problem of the coagulation in internally oxidized alloys; reference is made to an article published a few years ago<sup>17)</sup>. One can clearly observe oxide particles in internally oxidized Ag-Cd and Cu-Zn alloys (1 at. % alloy component) under

the optical microscope at a magnification of  $500\times$  (fig. 6); in (1 at. %) Ag-Zn alloys this was not possible. In most cases the hardness has already decreased considerably before the coagulation has proceeded so far that the particles are microscopically visible.



Fig. 6. An etched section through an internally oxidized Cu-Zn alloy (99% Cu-1% Zn) clearly shows oxide particles, when enlarged 500 times. The material was heated for  $3\frac{1}{2}$  hours at 950 °C in air.

A property of alloys which is very sensitive to coagulation is the electrical resistivity. In the case of alloys hardened by precipitation, the resistance mostly rises first with increasing coagulation, after which it falls below the original value. The resistivity maximum is reached sooner than the hardness maximum. In some cases we found an electrical resistivity for internally oxidized silver alloys which was considerably larger than for non-oxidized alloys; in the case of copper alloys hardened by oxidation, no increase in resistance is found. This shows that the coagulation in copper alloys proceeds much further than in silver alloys because Cu has more affinity for O than Ag.

The coagulation also causes the deformation of "inner oxide films" in some alloys containing rather strong concentrations of base elements (fig. 7). These more or less coherent layers of oxide, roughly parallel to the metal surface, may impede the internal oxidation in the deeper parts of the metal<sup>18)</sup>.

The hardness obtained by oxidation hardening is less sensitive to subsequent treatment at high temperatures than that obtained by cold working or precipitation hardening, which is a definite advantage. The occurrence of intercrystalline brittleness on the other hand is a disadvantage.

<sup>17)</sup> J. L. Meijering, Hardening of metals by internal oxidation, "Strength of Solids", Bristol Report, 140-150, 1948.

<sup>18)</sup> J. L. Meijering, Pittsburgh International Conference on surface reactions, 103, 1948.





Fig. 7. Inner oxide films formed in copper with 5 at. % beryllium. The films run roughly parallel to the metal surface, the arrow indicates the direction of the oxygen diffusion.

Apart from hardening by internal oxidation, hardening by internal nitriding, viz. nitride

hardening of steel is known. If steel with a small content of e.g. aluminium or chromium is heated at approximately 500 °C in ammonia gas, a very strong increase in hardness is observed. This phenomenon may be explained by analogy with the picture given above of oxidation hardness, except that nitrides (of Al or Cr) will occur instead of oxides.

**Summary.** The hardness of a metal is its resistance against plastic deformation. This resistance may vary widely for the various kinds of metal; some factors influencing it are the crystal structure (both distorted and undistorted), the presence of foreign atoms in the lattice, or the presence of foreign molecules or groups of atoms. The hardness of a metal may be affected in various ways: by heat treatment either with or without chemical reactions, by deformation (work-hardening), or by a special method of producing the metal (electrolytical deposition). Thermal methods of hardening after the forming, such as quench hardening, precipitation hardening and oxidation hardening, are discussed and these methods are illustrated by examples, the last, new, method somewhat more extensively.

## ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS' GLOEILAMPENFABRIEKEN

Reprints of these papers not marked with an asterisk \* can be obtained free of charge upon application to the Administration of the Research Laboratory, Kastanjelaan, Eindhoven, Netherlands.

**2005:** L. A. Ae. Sluyterman: Reactions of polypeptide esters in the solid state (thesis; Centen, Amsterdam 1951; 36 pp., 9 figs.).

Discussion of the behaviour of the methyl esters of the tripeptides diglycyl-glycine and alanyl-glycyl-glycine and of the tetrapeptides triglycyl-glycine and alenyl-diglycyl-glycine. It is shown that the tripeptides and the tetrapeptides react differently on heating in the solid state. The products obtained on heating have been investigated in a number of cases. In the discussion attention is drawn to the fact that the occurrence of an induction period and the migration of methyl groups as observed in some cases, and which preponderates above polycondensation, may be thought to be due to the rigidity of the crystal lattice.

**2006:** H. Bruining: Opneembuizen voor televisie, I. Indeling en wijze van werken van opneembuizen (T. Ned. Radiogenootschap **16**, 209-225, 1951, No. 5). (Television camera tubes, I. Types and mode of operation; in Dutch).

In this paper a general outline is given on different types of television camera tubes. After an introduc-

tion on the two ways of potential stabilisation of a surface of insulating material by scanning electrons of low and high velocity respectively, the constructions of the orthicon, image-orthicon, vidicon, iconoscope and image-iconoscope are discussed. The mechanisms of storage, measuring and wiping out of electric charges due to bombardment by photo-electrons or due to photoconductivity are treated in detail. At the end a survey is given of tubes with their special names (see Philips tech. Rev. **13**, 119-133, 1951, No. 5).

**2007:** P. Schagen: Opneembuizen voor televisie, II. Eigenschappen en onderlinge vergelijking van de opneembuizen (T. Ned. Radiogenootschap **16**, 227-242, 1951, No. 5). (Television camera tubes, II. Properties and inter-comparison of camera tubes; in Dutch).

A comparison is given of the performances of the four most important types of camera tube. The special tubes considered are: Tubes with low-velocity target stabilization: Orthicon (C.P.S.-emitter). Image orthicon (American type 5826). Tubes with high-velocity target stabilization: Iconoscope,



Image iconoscope (Philips type 5854). The following characteristics are discussed: (1) Characteristics connected with the method of target stabilization employed: a) linearity and gamma, b) effective exposure-time, c) spurious-signal generation, d) simplicity of installation and operation. (2) Some other characteristics: a) Signal-to-noise ratio, b) spectral response, c) stability, d) depth of focus, e) possibility of black-level restoration. (3) Sensitivity: for normal broadcasting purposes, where a good signal-to-noise ratio and reasonable depth of focus are required, the sensitivities of the tubes are found to be related as follows: Iconoscope: Image iconoscope: Orthicon: Image orthicon = 1 : 75 : 25 : 325 (see No. 2006).

**R 177:** W. Ch. van Geel and B. C. Bouma: La déformation des redresseurs électrolytiques et de quelques redresseurs à couche d'arrêt (Philips Res. Rep. 6, 401-424, 1951, No. 6). (The de-formation of electrolytic rectifiers and of some barrier layer rectifiers; in French.)

Electrolytic rectifiers and some other kinds of rectifiers can lose their rectifying properties by a de-formation process. By "deformation" we mean the act of sending a current through the system in the direction opposite to the current used during the formation of the oxide layer (blocking-layer). During the deformation the current passes in the direction of easy transmission. After deformation the system can be restored by sending a current through the system in the direction that was originally the blocking direction (re-forming). After re-forming the system recovers its rectifying properties. For re-forming also alternating voltage can be used. The rectifying properties can also be taken away and given back to synthetic rectifiers, e.g. Zr-ZrO<sub>2</sub>-(CuJ+J). The same can be done with rectifiers with a resin barrier layer, where a resin layer is brought onto a semiconductor and covered with a metal alloy and also with the ordinary type of selenium rectifier, where an alloy of low melting-point is brought directly onto the selenium layer. It is suggested that in all these cases rectification occurs because the barrier layer is built up in two parts which owe their conductivity to charge carriers of different sign (electron and hole semiconductors). By deformation one of the part loses its own character of conductivity

and there remain two parts with the same sign of conductivity and therefore rectification disappears.

**R 178:** Th. P. J. Botden: Transfer and transport of energy by resonance processes in luminescent solids (Philips Res. Rep. 6, 425-473, 1951, No. 6).

Transfer and transport of energy have been investigated in solids systems in which excitation does not produce free electrons and holes, so that these processes, if they do take place, are brought about by resonance. A distinction is made between those systems in which transport of energy may occur via the base lattice (described in part I) and the so-called sensitized systems in which transport may occur only via particular centres of the lattice (described in part II). Part I deals with the investigations on alkaline-earth tungstates and molybdates activated with samarium. Upon excitation in the absorption band of the base lattice with ultraviolet radiation both the tungstate or molybdate and the samarium fluorescence are emitted at low temperatures (below about 250 °K). At these temperatures the quantum efficiency of the samarium fluorescence is independent of temperature while for excitation by certain wavelengths the quantum efficiency of the tungstate or molybdate fluorescence is temperature-dependent, because the absorption changes with temperature. The excitation energy that leads to samarium fluorescence at low temperatures is mainly absorbed in tungstate or molybdate groups in the neighbourhood of the samarium ions and the energy is transferred from these groups to samarium. At the temperature at which the tungstate or molybdate fluorescence is quenched, or at a slightly higher temperature, the quantum efficiency of the samarium fluorescence increases with temperature due to the transfer of energy from tungstate or molybdate groups at larger distances from the samarium ions to these ions. This energy is transported from the groups first excited to groups in the neighbourhood of the samarium ions. The mechanisms of transfer and transport of energy are discussed. Calculations have been carried out on the basis of the theories of Mott-Seitz and of Möglich-Rompe for the quenching of fluorescence. The results are in favour of the theory of Mott-Seitz, but do not confirm that of Möglich-Rompe.